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(DOKLADY)

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NOUVELLE SÉRIE

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LES REPRÉSENTATIONS QUASI-CONFORMES ET LEURS SYSTÈMES DÉRIVÉS

Par M. LAVRENTIEV, de l'Académie des Sciences d'Ukraine

Dans la présente Note j'introduis la notion de «systèmes dérivés» pour les classes les plus générales de représentations quasi-conformes correspondant aux systèmes fortement elliptiques. Les propriétés des «systèmes dérivés» que nous considérons dans la suite permettent d'établir quelques propriétés nouvelles des solutions des systèmes d'équations différentielles.

1. Définitions. Nous dirons qu'un système d'équations

$$\Phi_1\left(x,\ y;\ u,\ v;\ \frac{\partial u}{\partial x},\ \frac{\partial u}{\partial y},\ \frac{\partial v}{\partial x},\ \frac{\partial v}{\partial y}\right)=0 \\ \Phi_2\left(x,\ y;\ u,\ v;\ \frac{\partial u}{\partial x},\ \frac{\partial u}{\partial y},\ \frac{\partial v}{\partial x},\ \frac{\partial v}{\partial y}\right)=0 \quad \Bigg\} \tag{1}$$

admet une représentation quasi-conforme d'un domaine D situé dans le plan $x,\ y$ sur un domaine Δ du plan $u,\ v,$ s'il existe une représentation homéomorphe du domaine D sur Δ réalisé par des fonctions

$$u = u\,(x,\ y) \\ v = v\,(x,\ y) \\ \Delta \doteq \begin{vmatrix} \dfrac{\partial u}{\partial x} & \dfrac{\partial u}{\partial y} \\ \dfrac{\partial v}{\partial x} & \dfrac{\partial v}{\partial y} \end{vmatrix} > 0 \quad \Bigg\} \tag{2}$$

vérifiant les équations (1). Nous dirons aussi que la représentation (2) correspond au système (1).

Soit (2) une représentation quasi-conforme qui correspond au système (1). Un couple de points correspondants étant donné, considérons la partie principale linéaire de la représentation

$$u - u_0 = u_x\,(x - x_0) + u_y\,(y - y_0) \\ v - v_0 = v_x\,(x - x_0) + v_y\,(y - y_0) \quad \Big\} \tag{3}$$

Considérons dans le plan $u,\ v$ le carré de côté égal à 1 ayant pour sommet le point $w_0 = u_0 + iv_0$ et pour côtés $\overline{w_0 w_1},\ \overline{w_0 w_2}$

$$w_2 - w_0 = (w_1 - w_0)\,e^{i\frac{\pi}{2}}$$

Désignons par ν l'angle du vecteur $\overline{w_0 w_1}$ avec l'axe des u

$$w_1 - w_0 = e^{i\nu}$$

Dans la représentation (3) il correspond à ce carré un certain parallélipipède Π_v; soient z_1 et z_2 les points qui correspondent à w_1, w_2.
Posons

$$z_1 - z_0 = V_v e^{i\alpha_v}$$

$$\theta_v = \arg \frac{z_2 - z_0}{z_1 - z_0}$$

$$W_v V_v \Delta = 1$$

où Δ est le déterminant de la transformation (3). Les quantités introduites (quel que soit v fixe) V_v, α_v, θ_v, W_v déterminent complètement le parallélipipède Π_v et peuvent être exprimées d'une manière élémentaire au moyen des coefficients de la transformation (3); les quantités seront nommées caractéristiques de la représentation. Les relations (1) peùvent être remplacées par deux relations entre les caractéristiques

$$\left. \begin{array}{l} W_v = F_1^{(v)}(V_v, \alpha_v; \ x, y; \ u, v) \\ \theta_v = F_2^{(v)}(V_v, \alpha_v; \ x, \ y; u, \ v) \end{array} \right\} \tag{4}$$

Nous dirons que le système (1) est f o r t e m e n t   e l l i p t i q u e si, quel que soit le nombre v, la représentation (4) de notre système vérifie les conditions suivantes: 1) les fonctions F_1 et F_2 sont univoques et continues pour toutes les valeurs des arguments; 2) il existe une constante positive k telle que pour toutes les valeurs des arguments on a

$$\kappa < \theta_v < \pi - k$$

3) quels que soient α_v, x, y, u, v fixes, la fonction F_1 est croissante par rapport à V_v, $V_v > 0$: $\partial F_1 / \partial V > k > 0$.
Remarquons que si le système (1) est homogène et linéaire par rapport aux dérivées partielles, les conditions pour qu'il soit elliptique au sens classique sont équivalentes aux conditions citées pour qu'il soit fortement elliptique.
Les définitions précédentes sont données dans une Note qui est sous presse dans les Comptes Rendue de l'Académie Ukrainienne. Dans la même Note j'indique certaines propriétés des représentations quasi-conformes et du théorème d'existence pour le problème de Riemann de la représentation des domaines plans: le problème de la représentation quasi-conforme correspondant au système fortement elliptique est résoluble si les fonctions F_1 et F_2 et les frontières des domaines vérifient certaines relations de la théorie des fonctions.
2. S y s t è m e s   d é r i v é s. Soit (2) la représentation quasi-conforme du domaine D sur le domaine Δ correspondant au système (4). Considérons dans le plan u, v un carré de côtés infiniment petits parallèles aux axes de coordonnées, et soit q le quadrilatère infiniment petit qui correspond à ce carré. Formons les rapports des côtés de q qui correspondent aux côtés du carré parallèles à l'axe u et à l'axe v. En rejetant dans les calculs les infiniment petits d'ordres supérieurs et en posant $P = \log V$ nous aurons

$$\left. \begin{array}{l} \dfrac{\partial P}{\partial v} = a_1 \dfrac{\partial P}{\partial u} + a_2 \dfrac{\partial \alpha}{\partial u} + a_3 \\[2mm] \dfrac{\partial \alpha}{\partial v} = b_1 \dfrac{\partial P}{\partial u} + b_2 \dfrac{\partial \alpha}{\partial u} + b_3 \end{array} \right\} \tag{5}$$

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où les coefficients a et b sont des fonctions connues de x, y, u, v, p et α:

$$a_1 = \frac{\partial W}{\partial V} \operatorname{cotg} \theta - \frac{\partial \theta}{\partial V} \frac{W}{\sin^2 \theta}$$

$$a_2 = \frac{1}{V} \left\{ \frac{\partial W}{\partial \alpha} \operatorname{cotg} \theta - \frac{\partial \theta}{\partial \alpha} \frac{W}{\sin^2 \theta} - W \right\}$$

$$a_3 = \operatorname{cotg} \theta \left\{ \frac{\partial W}{\partial u} \frac{1}{V} + \frac{\partial W}{\partial x} \cos \alpha + \frac{\partial W}{\partial y} \sin \alpha \right\} -$$

$$- \frac{W}{\sin^2 \theta} \left\{ \frac{\partial \theta}{\partial u} \frac{1}{V} + \frac{\partial W}{\partial x} \cos \alpha + \frac{\partial W}{\partial y} \sin? \right\}$$

$$b_1 = \frac{\partial W}{\partial V}$$

$$b_2 = \frac{1}{V} \left\{ \frac{\partial W}{\partial \alpha} + W \operatorname{cotg} \theta \right\}$$

$$b_3 = \frac{\partial W}{\partial u} \frac{1}{V} + \frac{\partial W}{\partial x} \cos \alpha + \frac{\partial W}{\partial y} \sin \alpha$$

Le système d'équations (5) sera nommé système dérivé de la représentation (2) considérée ou bien du système (4).

Théorème. *Si le système* (2) *est fortement elliptique, le système* (5) *est aussi elliptique*

$$- 4a_2 b_1 > (b_2 - a_1)^2$$

Signalons quelques cas particuliers. Si dans le système (2) W et θ ne dépendent pas explicitement des coordonnées, on aura $a_3 = b_3 = 0$, le système dérivé (5) sera un système homogène du type elliptique, les fonctions $P = P(u, v)$ et $\alpha = \alpha(u, v)$ réalisent la représentation quasi-conforme du domaine Δ sur une certaine surface de Riemann du plan P, α. Si P, α sont considérés comme variables indépendantes dans notre système homogène, ce système devient évidemment l i n é a i r e . Dans le cas des équations de la mécanique des gaz le système dérivé coïncide avec le système connu d'équations obtenu par S. A. Tchaplyguine par une méthode analytique.

Revenons au cas général. En considérant la partie homogène du système (5)

$$\left. \begin{array}{l} \dfrac{\partial P}{\partial v} = a_1 \dfrac{\partial P}{\partial u} + a_2 \dfrac{\partial \alpha}{\partial u} \\[2mm] \dfrac{\partial \alpha}{\partial v} = b_1 \dfrac{\partial P}{\partial u} + b_2 \dfrac{\partial \alpha}{\partial u} \end{array} \right\} \tag{5'}$$

comme un système qui correspond à une certaine représentation quasi-conforme, nous pouvons définir pour (5′) un système dérivé qui sera dit système dérivé du second ordre pour le système (4). Ainsi e n s u p p o s a n t q u e l e s f o n c t i o n s F_1, F_2, u, v a d m e t t e n t u n n o m b r e s u f f i s a n t d e d é r i v é e s , nous pouvons construire des systèmes dérivés d'ordre quelconque.

3. A p p l i c a t i o n s. En s'appuyant sur la notion de système dérivé et sur quelques propriétés des représentations quasi-conformes correspondant aux systèmes linéaires (1), on peut établir certains théorèmes concernant le lien entre la dérivabilité des fonctions F_1, F_2 et la dérivabilité des fonctions u, v. Indiquons l'un de ces théorèmes.

Théorème. *Si le système* (4) *est fortement elliptique et si les fonctions* F_1, F_2 *possèdent des dérivées partielles vérifiant la condition de Hölder, alors dans la représentation correspondante* (2) *des domaines bornés* D *et* Δ *les fonctions* P, α *possèdent des dérivées partielles vérifiant la condition de Hölder.*

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MATHEMATICS

ON A GENERALIZATION OF THE FEJER-RIESZ THEOREM

By B. J. LEVIN

(Communicated by S. N. Bernstein, Member of the Academy, 7. III. 1946)

The present note is devoted to a generalization of the well-known theorem of Fejer and Riesz [1]. Every non-negative trigonometric polynomial

$$T(x) = \sum_{-n}^{n} c_k e^{ikx} \quad (f(x) \geqslant 0, \ -\infty < x < \infty) \ .$$

may be represented in the form

$$T(x) = \Big| \sum_{0}^{n} b_k e^{ikx} \Big|^2 \quad (-\infty < x < \infty)$$

We shall consider instead of trigonometrical polynomials functions $f(x)$ of the form

$$f(x) = \int_{-A}^{A} e^{ixt} d\sigma(t) \quad (-\infty < x < \infty) \tag{1}$$

where $\sigma(t)$ $(\sigma(-A) = 0)$ is a certain complex-valued function of bounded variation in the interval $(-A, A)$. Using the decomposition $\sigma(t) = \sigma_a(t) + \sigma_s(t) + \sigma_d(t)$, $(-A \leqslant t \leqslant A; \ \sigma_a(-A) = \sigma_s(-A) = \sigma_d(-A) = 0)$, where $\sigma_a(t)$ and $\sigma_s(t)$ are, respectively, the absolutely continuous and the singular components of the function $\sigma(t)$, and $\sigma_d(t)$ is the function of jumps, we may now formulate our main result in the following manner.

Theorem 1. *Let*

$$f(x) \geqslant c > 0 \ (-\infty < x < \infty) \tag{2}$$

$$\inf_{x} \Big| \int_{-A}^{A} e^{ixt} d\sigma_d(t) \Big| > \int_{-A}^{A} |d\sigma_s(t)| \tag{3}$$

Then

$$f(x) = \Big| \int_{-A/2}^{A/2} e^{ixt} d\tau(t) \Big|^2 \quad (-\infty < x < \infty)$$

where $\tau(t)$ is a certain complex-valued function of bounded variation in the interval $(-A/2, A/2)$.

Observe at once that if $\sigma_s(t) \equiv 0$, then the condition (3) will be a corrolary of condition (2). From our proof of the theorem will also follow that if $\sigma_s(t) = 0$ or $\sigma_a(t) = 0$, then we may put, respectively, $\tau_s(t) = 0$ or $\tau_a(t) = 0$. Moreover, if $\sigma(t) = \sigma_d(t)$, then we shall also have $\tau(t) = \tau_d(t)$.

Lemma 1. *In order that the function*

$$f(z) = \int_{-\infty}^{\infty} e^{izt} d\sigma(t) \quad (\mathrm{var}\ \sigma(t) < \infty) \tag{4}$$

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should be an integral function of the exponential type A, it is necessary and sufficient that $d\sigma(t) = 0$ *for* $|t| > A$, *i. e.*

$$f(z) = \int_{-A}^{A} e^{izt} d\sigma(t) \tag{5}$$

L e m m a 2. *Let* $f(z)$ *be an integral function-of exponential type* A *bounded on the real axis. Then*

$$-\sum_{k=1}^{\infty} \operatorname{Im}\left(\frac{1}{a_k}\right) < \infty \tag{6}$$

Herein denote the a_k $(k = 1, 2, \ldots)$ the roots of the function $f(z)$ situated in the upper half-plane.

P r o o f. From the Phragmen-Lindelöff principle easily follows where M

$$\left| e^{iA} f(z) \prod_{k=1}^{n} \frac{z - \bar{a}_k}{z - a_k} \right| < M \quad (\operatorname{Im} z > 0) \tag{7}$$

is a constant not depending on n. Hence immediately follows the convergence of the infinite product and the series

$$\prod_{k=1}^{\infty} \left| \frac{i - a_k}{i - \bar{a}_k} \right|^2, \ \sum_{k=1}^{\infty} \left(1 - \left| \frac{i - a_k}{i - \bar{a}_k} \right|^2 \right) = 4 \sum_{k=1}^{\infty} \frac{\operatorname{Im} a_k}{|i - a_k|^2} \tag{8}$$

and, consequently, also of the series (6).

L e m m a 3. *Let* $f(z)$ *be an integral function of exponential type* A *positive and bounded on the real axis, and let*

$$f(z) = c e^{kz} \prod_{k=1}^{\infty} \left(1 - \frac{z}{a_k} \right) e^{\frac{z}{a_k}} \tag{9}$$

Then the function

$$\Phi(z) = \sqrt{c} \, e^{-i\Delta z + \frac{k}{2} z} \prod_{k=1}^{\infty} \left(1 - \frac{z}{a_k} \right) e^{\frac{z}{a_k}} \quad \left[\Delta = \sum_{k=1}^{\infty} \operatorname{Im}\left(\frac{1}{a_k}\right) \right]$$

where the z_k *are the roots of* $f(z)$ *with negative imaginary parts, is an integral function of exponential type* $A/2$[*].

P r o o f. Let

$$\tilde{\Phi}(z) = \prod_{k=1}^{\infty} \left(1 - \frac{z}{|a_k|} \right) e^{\frac{z}{|a_k|}}$$

Then

$$\ln \left| \frac{\Phi(ir)}{\tilde{\Phi}(ir)} \right| = \sum_{k=1}^{\infty} \ln \left| 1 + \frac{ir\left(\frac{1}{|a_k|} - \frac{1}{a_k} \right)}{1 - \frac{ir}{|a_k|}} \right| <$$

$$< 2r \sum_{k=1}^{\infty} \frac{\operatorname{Im}\left(\frac{1}{a_k} \right)}{\left| 1 - \frac{ir}{|a_k|} \right|} < \varepsilon r \quad (\varepsilon > 0, \ s > Rs)$$

_____

[*] In the sequel we use only the fact that $f(x) = |\Phi(x)|^2$ where $\Phi(x)$ is an integral function of exponential type $A/2$ not vanishing in the upper half-plane. In this form Lemma 3 is to be found in the papers of Krein ([2], [3]).

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On the other hand, denoting by $n(t)$ the number of roots of the function $f(z)$, the modulus of which does not exceed t, we shall have

$$\ln |\widetilde{\Phi}(ir)|^2 = \int_0^\infty \ln \left| 1 + \frac{r^2}{t^2} \right| dn(t) = 2r^2 \int_0^\infty \frac{tN(t)\,dt}{t^2+r^2} \left(N(t) = \int_0^t \frac{n(r)}{r} dr \right)$$

By Jensen's theorem we find

$$N(t) = \frac{1}{2\pi} \int_0^{2\pi} \ln |f(te^{i\varphi})| \, d\varphi \leqslant \frac{2A}{\pi} t + \ln M$$

whence

$$|\widetilde{\Phi}(ir)| \leqslant \sqrt{M} e^{\frac{A}{2} r}, \quad |\Phi(ir)| \leqslant K e^{\left(\frac{A}{2}+\varepsilon\right) r}$$

The function $\Phi(z)$ is of the first genus, increases exponentially along the imaginary axis and is bounded on the real axis by the number $\sqrt{M}$. Hence easily follows that

$$|\Phi(z)| \leqslant \sqrt{M} e^{\frac{A}{2} |y|}$$

A similar inequality holds for the function $\overline{\Phi}(z)$ the roots of which are conjugated to the roots of $\Phi(z)$, and since the function $\Phi(z) \times \overline{\Phi}(z) = f(z)$ is of the type A, the type of $\Phi(z)$ is equal to $A/2$.

Proceeding now to the proof of the theorem, we observe in the first place that by a well-known theorem of N. Wiener[4] from the conditions (2) and (3) it follows that the function $\ln f(x)$ is representable by the Fourier-Stieltjes integral

$$\ln f(x) = \int_{-\infty}^\infty e^{ixt} d\sigma_2(t) \quad (\text{var } \sigma_2(t) < \infty)$$

or, in virtue of the reality of $\ln f(x)$

$$\frac{1}{2} \ln f(x) = \text{Re} \left[\int_0^\infty e^{ixt} d\sigma_2(t) \right]$$

The real parts of the functions $\ln \Phi(z) e^{i \frac{A}{2} z}$ and $\psi(z) = \int_0^\infty e^{izt} d\sigma_2(t)$ evidently coincide on the real axis. Besides, both functions are regular and bounded in the upper half-plane (the boundedness of the function $\Phi(z) e^{i(A/2)z}$ follows from the Phragmen-Lindelöf's principle). The difference between these functions can be continued into the lower half-plane and so, by Liouville's theorem, must be an imaginary constant. Thus

$$\ln \Phi(z) e^{i \frac{A}{2} z} = \int_0^\infty e^{izt} d\sigma_2(t) + i\mu$$

It is evident that an integral function of a function of the class (4) also belongs to this class, and consequently

$$\Phi(z) = e^{i \frac{A}{2} z} e^{\psi(z) + i\mu} = \int_{-\infty}^\infty e^{izt} d\sigma_1(t) \quad (\text{var } \sigma_1(t) < \infty)$$

In virtue of Lemmas 1 and 3 hence follows

$$\Phi(z) = \int_{-A/2}^{A/2} e^{izt} d\sigma(t) \tag{10}$$

and the theorem is proved.

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Remark. If in (1) $\sigma(t)$ is a function of jumps, then applying N. Wiener's theorem on absolutely convergent trigonometrical series[5], we obtain (10) where $\sigma_1(t)$ will be also a function of jumps.

Observe also that this theorem cannot be strengthened for generalized trigonometrical polynomials, *i. e.* that there exist positive polynomials which cannot be represented as the square of the modulus of a polynomial.

Generalizing a theorem of M. Krein *, A. Artemenko established the following proposition:

Theorem 2 (Artemenko). *An arbitrary Hermite-positive on the interval $(-A, A)$ function $f(x)$ may be so defined for values of $x \bar\in (-A, A)$ that it will be Hermite-positive on the whole axis.*

Proof. From the definition of Hermite-positive functions directly follows: $f(-\lambda) = f(\lambda)$, $\overline{f(0)} \geqslant |f(\lambda)|$. Define on all functions of the form

$$\varphi(x) = \sum_{k=1}^{\infty} c_k e^{i\lambda_k x} \left(\sum_{k=1}^{\infty} |c_k| < \infty, \ |\lambda_k| \leqslant A \right)$$

the functional

$$F[\varphi] = \sum_{k=1}^{\infty} c_k f(\lambda_k)$$

If $\varphi(x) \geqslant m > 0$, then, by Theorem 1,

$$\varphi(x) = \left| \sum_{k=1}^{\infty} a_k e^{i\lambda_k x} \right|^2$$

and, consequently,

$$F[\varphi] = F\left[\left| \sum_{k=1}^{\infty} a_k e^{i\lambda_k x} \right|^2 \right] = \sum_{k,j=1}^{\infty} f(\lambda_k - \lambda_j) a_k \bar{a}_j \geqslant 0$$

If $\varphi(x) \geqslant 0$, then for arbitrary $\gamma > 0$ we have $F[\varphi] + \gamma f(0) = = F[\varphi(x) + \gamma] \geqslant 0$, and, consequently, $\gamma > 0$ being arbitrary, $F[\varphi] \geqslant 0$.

Thus, $F[\varphi]$ is a positive functional, which by a well-known theorem (cf. for instance[5], p. 154) may be continued as a positive functional to the set of all almost periodic functions with an absolutely convergent Fourier series. The function $\tilde{f}(\lambda) = F[e^{i\lambda x}]$, $(-\infty < \lambda < \infty)$, yields the required supplementary definition of the function $f(\lambda)$.

The method which we used in the proof of Theorem 2 is due to A. Artemenko, who has applied it to a problem closely related to that considered in the present note. But not possessing Theorem 1, he could not apply it to the proof of Theorem 2, for which he gives another proof in his unpublished thesis.

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* M. Krein has established Theorem 2 (cf. [6], Theorem 1) for continuous Hermite-positive functions.

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ON A SPECIAL FAMILY OF INFINITE UNITARY MATRICES

By N. P. ROMANOV

(Communicated by I. M. Vinogradow, Member of the Academy, 20. XII. 1945)

In my paper [1] I prove incidently the following proposition:

T h e o r e m. *The matrix* c_ω *with elements*

$$c_{ik} = \frac{\omega(k)}{\sqrt{\sigma\Omega(i)}} \sum_{d/(i,k)} \mu\left(\frac{1}{d}\right) |\omega(d)|^{-2}$$

where $\omega(n)$ *satisfies the conditions*

$$\omega(n) \neq 0, \quad \omega(ab) = \omega(a)\omega(b), \quad \sum_{n=1}^{\infty} |\omega(n)|^2 < \infty$$

and where

$$\sigma = \sum_{n=1}^{\infty} |\omega(n)|^2, \quad \Omega(n) = \sum_{d/n} \mu\left(\frac{n}{d}\right) |\omega(d)|^{-2} =$$

$$= |\omega(n)|^{-2} \prod_{p/n} (1 - |\omega(p)|^2)$$

—*is unitary, i. e.*

$$C_\omega^* C_\omega = C_\omega C_\omega^* = E$$

Herein denotes $E = \|\delta_{ik}\|$ the unit matrix and the operation * transforms an arbitrary matrix $A = \|a_{ik}\|$ into $A^* = |a_{ik}^*|$, where $a_{ik}^* = a_{ik}$.

It may be shown that the condition

$$\sum_{n=1}^{\infty} |\omega(n)|^2 < \infty$$

is equivalent to the conditions: 1) $|\omega(p)| < 1$ (p being an arbitrary prime); 2) $\sum_p |\omega(p)|^2 < \infty$ where p runs through all primes.

Since every multiplicative function is uniquely determined by its values for prime arguments, the matrix C_ω depends on the parameters $x_1, x_2, \ldots$, where $x_i = \omega(p_i)$, $p_1, p_2, \ldots$ being the sequence of all primes. Thus, $c_{ik} = c_{ik}(x_1, x_2, \ldots)$, where $x_1, x_2, \ldots$ are arbitrary complex numbers satisfying the conditions: 1) $0 < |x_i| < 1$ ($i = 1, 2, \ldots$); 2) $\sum_{i=1}^{\infty} |x_i^2| < \infty$.

This infinity-parametric family of infinite unitary matrices differs from the family of unitary matrices given by Cayley's method in that the latter family comprehends practically all unitary matrices, whereas the family considered here consists of matrices of a rather special character with specifically arithmetical properties. Besides, our method of const-

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ruction of unitary matrices has no finite-dimensional analogue, whereas the method of Cayley has such an analogue.

Putting $\omega(n) = |\omega(n)| e^{i\theta_n}$ we shall evidently have $C_\omega = M_\theta C_{|\omega(n)|}$ where the diagonal matrix M_θ has the elements $m_{ik} = \delta_{ik} e^{i\theta_k}$. Hence we see that it suffices to consider the family C_ω where $\omega(n) > 0$. In the sequel we shall therefore consider the parameters x_1, x_2, ... as positive. Then, as we have shown in (^{1})

$$\frac{\partial C_\omega}{\partial x_r} = C_\omega H_r, \qquad \frac{\partial C_\omega}{\partial x_r} = \tilde{H}_r C_\omega$$

where the matrix $H_r = \|h_{ik}^r\|$ is defined in the following manner: $h_{ik}^r = 0$ if the ratio i/k is not of the form p_r^λ, where p_r is the r-th prime in the sequence of all primes $p_1 < p_2 < p_3 < \dots$ and λ is a positive, or a negative integer, or zero; $h_{ik}^r = \mathrm{sgn}\,\lambda\; x_r^{|\lambda|-1}$ if $i/k = p_r^\lambda$. Similarly, $\tilde{H}_r = \|\tilde{h}_{ik}^r\|$ where $\tilde{h}_{ik}^r = 0$ if i/k is not of the form p_r^λ, and $\tilde{h}_{ik}^r = \mathrm{sgn}\,\lambda\; x_r^{|\lambda|-1}$ if $i/k = p_r^\lambda$ and, moreover, p_r/i, p_r/k. Further, $\tilde{h}_{ik}^r = \mathrm{sgn}\,\lambda \cdot x_r^{|\lambda|-1}/\sqrt{1 - x_r^2}$ if $i/k = p_r^\lambda$ and if at least one of the numbers i, k is not divisible by p_r. Thus, «the left and the right logarithmic derivatives» of the matrix C_ω by the parameter x_r contain a large number of zeros and depend only on x_r, which *a priori* is not at all obvious, but follows by rather involved computations.

From the preceding relations follows also

$$C_\omega H_r = \tilde{H}_r C_\omega, \quad \text{or} \quad \tilde{H}_r = C_\omega H_r C_\omega^{-1}$$

Putting $x_i = x_i(t)$ where the $x_i(t)$ are arbitrary differentiable functions satisfying the conditions $0 < x_i(t) < 1$, $\sum_{i=1}^\infty x_i(t)^2 < \infty$, we obtain an one-parametric family of orthogonal matrices $C_t = C_{\omega(n,\,t)}$ for which

$$\frac{dC_t}{dt} = \frac{\partial C_t}{\partial x_1}\frac{dx_1}{dt} + \frac{\partial C_t}{\partial x_2}\frac{dx_2}{dt} + \dots = C_t\left(H_1\frac{dx_1}{dt} + H_2\frac{dx_2}{dt} + \dots\right) = C_t H$$

and $dC_t/dt = \tilde{H}C_t$ where $\tilde{H} = \tilde{H}_1\frac{dx_1}{dt} + \tilde{H}_2\frac{dx_2}{dt} + \dots$

The matrix $H = \|h_{ik}(t)\|$ has the elements $h_{ik}(t) = \sum_{r=1}^\infty h_{ik}^r \cdot dx_r/dt = 0$ if i/k is not a power of a prime, and $h_{ik}(t) = \mathrm{sgn}\,\lambda\; x_s^{|\lambda|-1}(dx_s/dt)$ if $i/k = p^\lambda$, where λ is a positive, or a negative number, or zero, p is a prime and s is its number in the increasing sequence of all primes. In precisely the same way $\tilde{H} = \|\tilde{h}_{ik}(t)\|$ where $\tilde{h}_{ik}(t) = 0$ if $h_{ik}(t) = 0$, and $\tilde{h}_{ik}(t) = \mathrm{sgn}\,\lambda\; x_r^{|\lambda|-1}(dx_r/dt)$ if $i/k = p_r^\lambda$, $p_r/(i,k)$ and $\tilde{h}_{ik}(t) = \mathrm{sgn}\,\lambda \cdot (x^{|\lambda|-1}/\sqrt{1 - x_r^2}) dx_r/dt$ if $i/k = p_r^\lambda$, $p_r/(i,k)$.

From the relations $dC_t/dt = C_t H$ and $dC_t/dt = \tilde{H}C_t$ follows $C_t^*(dC_t/dt) = H$, $(dC_t/dt)C_t^* = \tilde{H}$, whence we obtain the following peculiar orthogonality relations: $\sum_{r=1}^\infty c_{ri}(dc_{rk}/dt) = 0$ if i/k is not an integral power of a prime, and $\sum_{r=1}^\infty c_{ir}(dc_{kr}/dt) = 0$ in the same way.

From the family under consideration it is not possible to choose any unitary group of the form e^{Bt}, where B is a constant Hermitian matrix, whatever be the choice of the functions $x_i(t)$, since the equality $C_t = e^{Bt}$ implies (being derived by t) $C_t^*(dC_t/dt) = (dC_t/dt)C_t^* = B$; at the same time,

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as we have seen, $C_t^* (dC_t/dt) = H$, $(dC_t/dt) C_t^* = \tilde{H}$, and H and $\tilde{H}$ are evidently equal neither to each other nor to constant matrices for any choice of the $x_i(t)$.

This fact alone reveals the rather special character of the family of matrices under consideration. I was led to this family of matrices starting from some problems of the theory of numbers, but I think that it deserves some attention also from the point of view of its analytic and algebraic structures.

<div align="right">Received
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THEORY OF ELASTICITY

BENDING OF A CYLINDRICAL SHELL UNDER A CONCENTRATED LOAD

By J. N. RABOTNOV

(Communicated by L. S. Leibenson, Member of the Academy, 19. XII. 1945)

The equations of the boundary zone, given in a previous paper by the author [1], make it possible to consider the problem of a shell subjected to the action of a concentrated force normal to its surface. In the case of cylindrical and spherical shells the problem reduces to the solution of equations of the second degree characterized by some peculiar features. The spherical shell was discussed by Geckeler [2], the case of a cylindrical shell is solved in this paper. It should be noted that Geckeler is not consistent enough in disregarding small terms, and as a result his solution for the deflection contains a part that may become infinite. This is not compatible with the well-known result for a plate. Our equations corroborate Geckeler's result in its basic part and at the same time do not lead to the above-mentioned error.

For a cylindrical shell the basic equations may be written as follows:

$$B \Delta\Delta\varphi = \frac{1}{a} \frac{\partial^2 F}{\partial u^2}, \quad \Delta\Delta F = -C(1-\mu^2)\frac{1}{a}\frac{\partial^2 \varphi}{\partial u^2} \tag{1}$$

where a is the shell radius, u is the coordinate measured along the generatrix.

Substituting the variables $\xi = u/a$, $\eta = v/a$ and excluding F, we obtain

$$\Delta\Delta\Delta\Delta\varphi = -\nu^4 \frac{\partial^4 \varphi}{\partial \xi^4} \tag{2}$$

where

$$\nu^4 = \frac{3(1-\mu^2)a^2}{h^2}$$

Let us consider the following equation of the second degree

$$\Delta X = \varepsilon\nu\sqrt{i}\,\frac{\partial X}{\partial \xi} \tag{3}$$

If its solution is $X = X_1 + iX_2$ when $\varepsilon = +1$, and $X = X_3 + iX_4$ when $\varepsilon = -1$, then the functions X_1, X_2, X_3, X_4 satisfy equation (2) and its general solution is

$$\varphi = C_1 X_1 + C_2 X_2 + C_3 X_3 + C_4 X_4$$

Putting in equation (3)

$$X = Z \exp\left(\varepsilon\frac{\nu\sqrt{i}}{2}\xi\right)$$

where Z is a function of $\rho = \sqrt{\xi^2 + \eta^2}$, we can transform it into Bessel equation

$$\frac{d^2 Z}{d\rho^2} + \frac{1}{\rho}\frac{dZ}{d\rho} - \frac{\nu^2 i}{2}Z = 0$$

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Let

$$Z = K_* \left(\frac{\nu \sqrt{i}}{\sqrt{2}} \rho \right) = \ker \frac{\nu \rho}{\sqrt{2}} + i \, \kei \frac{\nu \rho}{\sqrt{2}}.$$

and construct the functions X_i.

The symmetrical combination

$$\varphi = -A \left\{ \ch \frac{\nu \xi}{2\sqrt{2}} \cos \frac{\nu \xi}{2\sqrt{2}} \kei \frac{\nu \rho}{\sqrt{2}} + \sh \frac{\nu \xi}{2\sqrt{2}} \sin \frac{\nu \xi}{2\sqrt{2}} \ker \frac{\nu \rho}{\sqrt{2}} \right. \tag{4}$$

solves the problem for a concentrated force, applied normally to the shell surface at the point $\xi = \eta = 0$.

Indeed, developing φ in the vicinity of the origin of coordinates we have

$$\varphi = A \frac{\nu^2}{32} (5\xi^2 + \eta^2) \ln \rho + \cdots$$

The shearing force on the arc of a geodetic circle of radius ρa having its centre at the origin, may be determined from the relation

$$N = N_1 \frac{d\xi}{ds} + N_2 \frac{d\eta}{ds}$$

where

$$N_1 = -\frac{B}{a^2} \frac{\partial}{\partial \xi} \Delta \varphi, \quad N_2 = \frac{B}{a^2} \frac{\partial}{\partial \eta} \Delta \varphi$$

After computing we obtain

$$N = \frac{B}{a^2} \frac{3\nu^2}{8} \frac{A}{\rho} \tag{5}$$
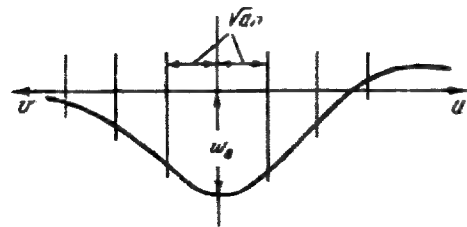
The condition of equilibrium leads to

$$A = \frac{4a^2 Q}{3 : B\nu^2} \tag{6}$$

The function φ, as may be easily shown, coincides with the normal displacement up to terms neglected in the approximate theory.

The maximum deflection by formula (4), when $\rho = 0$ is

$$\varphi_0 = \frac{Qa}{3B\nu^2} = \frac{Qa \sqrt{3(1-\mu^2)}}{6 \, E h^2} \tag{7}$$

Using asymptotic expansions for Bessel functions, we see that the solution found diminishes rapidly boith in the direction of coordinate ξ and in the direction of η.

For the stress function F at the origin we obtain a logarithmic peculiarity, the stress on the arc of an infinitesimal geodetic circle is directed along its radius and tends to infinity as $1/\rho^2$.

Taking this stress into account in the equation of statics results in it being necessary to introduce in the denominator of (6), instead of $3\nu^2$, the expression $3\nu^2 + 4$. However, this correction is outside of the limits of accuracy of the theory and should not be introduced. The figure is a graph of the function φ in the axial and transverse cross-sections of the shell.

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PHYSICS

ANGULAR DISTRIBUTION OF FAST ELECTRONS SCATTERED BY ALUMINIUM NUCLEI*

By L. ARZIMOVICH and I. PERRYMOND

(Communicated by A. F. Joffé, Member of the Academy, 2. II. 1946)

According to quantum-mechanical collision theory the probability of scattering of fast electrons by atomic nuclei of light elements must obey the following formula originally derived by Mott ([1])

$$I(\theta) = \frac{n}{4}\left(\frac{Ze^2}{mv^2}\right)^2\left(\frac{1}{\sin^4\frac{\theta}{2}} - \frac{\beta^2}{\sin^2\frac{\theta}{2}} + \frac{\pi\beta Z}{137}\frac{\cos^2\frac{\theta}{2}}{\sin^2\frac{\theta}{2}}\right)$$
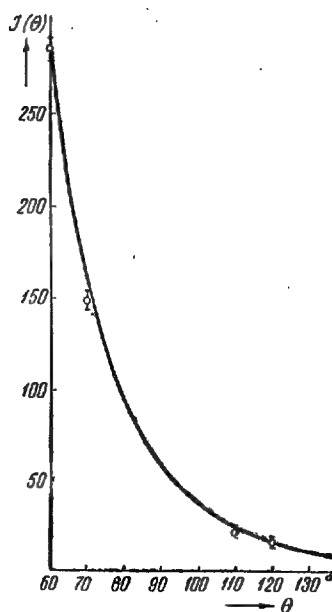
Fig. 1. Al; $E = 0.75$ MeV
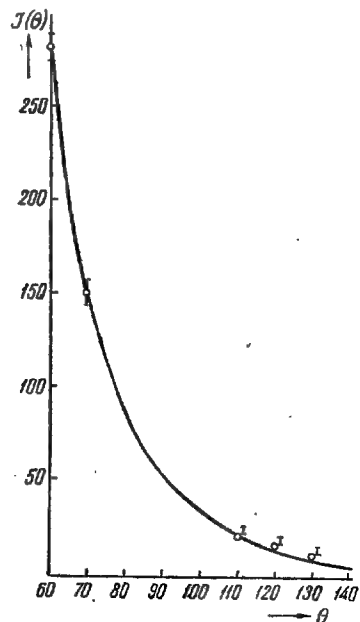
Fig. 2. Al; $E = 1$ MeV

where $I(\theta)$ is the relative number of electrons scattered per unit solid angle in the direction θ, n is the number of atoms per 1 cm² of the scattering layer; Z is the atomic number of the element; v is the electron velocity; m is the relativistic mass of an electron; e, its charge.

* This note was written in June 1941.

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For electrons with an energy of some 1—2 MeV the above formula has been checked experimentally by many workers. Most of them studied the scattering in a cloud chamber. The results of these investigations disagree with each other. Some authors found large deviations from Mott's formula in the angular distribution of electrons scattered by light substances.

In this paper are reported the results of an investigation on the angular distribution of electrons scattered by aluminium.

The method by which it was carried out seems better to fit the intended purpose compared to observations in a cloud chamber.

Radon preparations up to 200—500 millicuries were used as sources of high-speed electrons. A magnetic lens sorted out of the continuous β-spectrum a homogeneous electron beam with a maximum divergence of 16° (8° to both
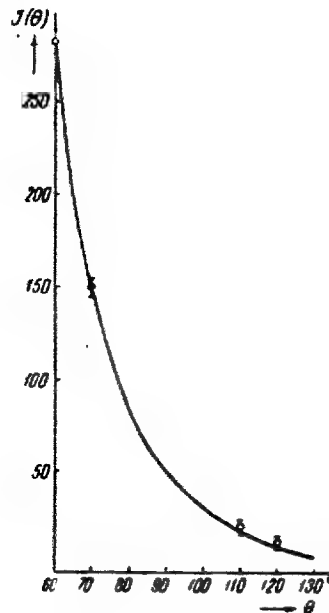
Fig. 3. Al; $E = 1.5$ MeV

sides of the axis). The relative width of the energy interval was about 15 per cent. The beam emerged from the monochromator through an opening 6 mm in diameter and was refocused on the scatterer by a second magnetic lens; the diameter of the focal spot was about 10 mm. This second focusing was necessary to prevent the scatterer from being attained by slow electrons. Moreover, this double focusing allowed the scatterer to be situated at a large distance from the radioactive source (150 cm), thus effecting a large reduction of the background arising from γ-rays.

The scatterers were mounted on a thin wire frame and placed inside a chamber, which was evacuated together with the magnetic spectrograph to a pressure of ~10⁻⁴ mm Hg. On the lateral cylindrical surface of the chamber a duralumin grid covered by a thin aluminium foil was mounted. The scattered electrons traversing this foil were recorded by a Geiger-Müller counter. The radius of the cylindrical chamber surface was 8.5 cm; this was also the distance from the scatterer to the counter.

Measurements were possible in an angular range of 60 to 140°. The solid angle subtended at the counter was 1.7×10^{-3}, and the angle range $\Delta\vartheta$ was 6°. The primary beam, after traversing the scatterer, passed into a glass tube 100 cm long and 18 cm in diameter. Owing to the presence of this tube the distorting influence of a backward scattering of the primary beam was

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eliminated. Moreover, to suppress the small backward scattering from the chamber wall an aluminium partition was inserted inside the chamber. The field of the second magnetic lens could not penetrate into the scattering chamber, as both lenses were screened by heavy iron shields.

The measurements were performed with electrons having energies 0.75, 1 and 1.5 MeV.

The scatterers were from 14 μ to 38 μ thick. The proportionality of the scattered intensity to the thickness of the scattering layer was carefully checked.

The results of the measurements, showing the angular distribution of electrons scattered by Al, are given in Figs. 1, 2 and 3. The scattering angle θ is plotted as abscissae, and the measured quantity $I(θ)$ in arbitrary units, as ordinates. The maximum statistical error at large scattering angles can attain $10 \div 15$ per cent. The full line curves are computed with Mott's formula. The experimental and theoretical values $I(θ)$ are made to coincide at $θ = 60°$.

From these curves, the exponential data obtained appear to be in agreement with Mott's formula.

Accurate measurement of the absolute value of $\dot{I}(θ)$ was not the aim of this investigation. Nevertheless, it follows from our measurements that in the investigated energy range the absolute value of the scattering probability cannot possibly differ from the theoretical value more than 1.5 or 2 times.

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CONCERNING MAXIMUM VALUES OF SOLAR RADIATION INTENSITY

By N. N. KALITIN

(Communicated by A. A. Grigoriev, Member of the Academy, 20.II.1946)

The values of solar radiation intensity, measured near the Earth's surface (not high above sea level), are dependent on the value of the solar constant, the variation of the distance between the Sun and the Earth and the decrease suffered by a Sun ray in the atmosphere.

The conditions of the actual atmosphere may occasionally be such as to reduce the decrease in the radiation to its minimum, whereas the intensity of solar radiation attains its maximum. This value may be considered as one of the characteristic features of the radiation climate of a given place. In order to obtain the actual maximum values, it is necessary to dispose of series of measurements of solar radiation intensity, carried out continuously during a number of years.

Three points of the USSR, Yakutsk, Tashkent, Pavlovsk (near Leningrad) located in absolutely different climatic conditions, have been chosen for the study of the maximum values of solar radiation. For Pavlovsk observational data were available for 27 years (1914—1940); for Yakutsk, for 13 years (1931—1943), and for Tashkent, for 18 years (1926—1943).

In the course of a year the maximum values are distributed as follows:

Maximum Values of Radiation (calories)

	I	II	III	IV	V	VI	VII	VIII	IX	X	XI	XII
Yakutsk	0.99	1.26	1.48	1.50	1.47	1.46	1.39	1.45	1.44	1.33	1.19	0.82
Pavlovsk	1.10	1.25	1.40	1.43	1.40	1.40	1.36	1.37	1.34	1.29	1.16	1.07
Tashkent	1.48	1.48	1.50	1.50	1.45	1.45	1.40	1.43	1.43	1.43	1.42	1.41

The table shows, first, that in spite of the difference in climatic conditions, the annual maximum is observed in one and the same month—April (in Tashkent, in March and April). This is explained by the fact that in April the Sun has already reached a considerable height, the absolute air humidity being not high. For Tashkent the annual course is no significant one, whereas for Yakutsk it is large, which is explained by the latitude of that place, and consequently by the height of the Sun.

For Yakutsk, Pavlovsk and Tashkent the maximum values are 1.50; 1.43; 1.50 cals, respectively. Solar radiation intensity for an ideal atmosphere can be computed for the periods when these values had been observed and compared then with observational data. The ratios of the latter to the «ideal» values are found to be equal to 0.95; 0.89; 0.90, respectively. As may be seen, the actual atmosphere may in some exceptional cases closely approach the

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ideal, especially in Yakutsk, where the Sun reaches as early as in April a considerable height, the absolute humidity is low and but little dust is contained in the air (snow cover).

Permanent observations had been carried out in Pavlovsk (with the aid of an actinograph) during 27 years, for which reason it is possible to undertake a detailed analysis of the conditions under which the maximum values of solar radiation intensity were obtained at this place.

Fig. 1 (bottom) shows the secular run of the maximum values for 27 years; as may clearly be seen, the curve consists of two distinct parts. Its lower portion, which corresponds to the mean value of radiation equal to 1.34 cals, embraces the period from 1914 to 1925; its upper portion corresponding to the years 1926—1940 shows a maximum of 1.39 cals.

This may be explained by the fact that from 1913 to 1923 the atmosphere was of a decreased transparency, apparently due to the eruption of the Katmai
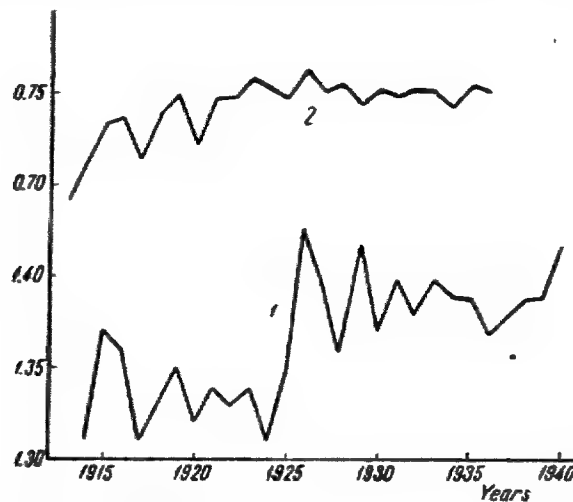
Fig. 1. *1*, Maximum values of solar radiation intensity:
2, coefficients of atmosphere transparency

Volcano, which occurred in the Alaska Peninsula on June 6th, 1912 ([1]). This conclusion permits to plot a curve of average year coefficients of atmosphere transparency for Pavlovsk ([2]), shown at the top of Fig. 1. As may be seen, the curve shows a rise, with certain fluctuations in either direction, from 0.690 for 1913 (transparency coefficient for 1912 equalling 0.570) and proceeds till 1924, where it stops rising, its further run being horizontal.

In order to make clear what meteorologic conditions are favourable for maximum transparency, values of maximum radiation were chosen for every month, and a study of the course of meteorologic elements was carried out for a period of 11 days: five days before the maximum and five days after it. Under investigation were the following conditions: absolute humidity, relative humidity, temperature and pressure (daily means) observed at Pavlovsk. For the 6th day (maximum solar radiation) the values of these elements were put equal to 0. Curves of Fig. 2 show the average fluctuations for every month of the values corresponding to the 6th day.

From Fig. 2 it is evident that the maximum intensity of solar radiation depends upon the passage over the place of observation of air masses of an increased pressure, decreased temperature, absolute and respective humidity. Such a course of meteorologic elements has been observed for each month, its expression varying as to the case.

The curve of relative humidity is the most interesting feature of the relation established. The maximum intensity of solar radiation depends not only

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upon the selective absorption of water vapour contained in the atmosphere (absolute humidity), but also upon the loss on dispersion due to water molecules condensed on the condensation cores (a definite condition of relative humidity). This conclusion is corroborated by certain studies of water vapour condensation in the atmosphere ([3]) and the decrease of solar radiation caused by it ([5]), carried out during the recent years.
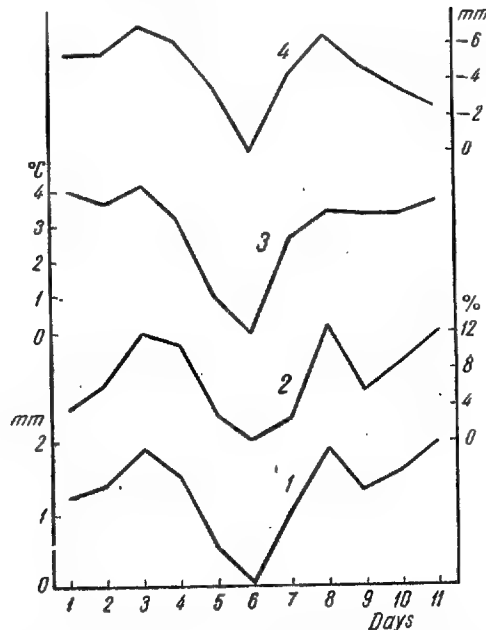
Fig. 2. Course of meteorologic elements at maximum solar radiation intensity. *1*, absolute humidity; *2*, relative humidity; *3*, temperature; *4*, pressure

There are scarcely any studies devoted to the maximum of solar radiation intensity. Gorczynski ([4]) points out that the most favourable conditions for the maximum are created by desert climate. Whether this is actually the case, we cannot say for lack of information available on solar radiation values observed under different climatic conditions. It is therefore highly desirable to undertake observations based upon long series of uniform actinometer measurements, secured under different climatic conditions.

All radiation values are given in Ångstrem units.

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GEOPHYSICS

A THEORY OF EQUATORIAL COUNTER-CURRENTS IN THE OCEAN

By W. STOCKMANN

Let us consider some region of the ocean, bounded by a rectangular contour and elongated in the zonal direction, *i. e.* along a parallel, which we take for the X-axis of a rectangular system of coordinates. We shall denote by L the length of the contour in this direction, while its length in the transverse direction coinciding with the Y-axis will be denoted by l $(L \gg l)$. We shall define the direction of the X-axis toward E to be positive, as likewise the direction of the Y-axis toward N.

If the water of the ocean is homogeneous (we assume its density to be equal to 1), we shall have, according to Eckman, the following expressions for the components of the total current S from the surface $(z = 0)$ to the ocean bed $(z = H)$ in the case of a steady wind and an inclined ocean level created by the piling-up effect of the wind

$$S_x = \int_0^H u\, dz = B\gamma_x + b\gamma_y + cT_y; \quad S_y = \int_0^H v\, dz = B\gamma_y - b\gamma_x - cT_x \quad (1)$$

Here u and v denote the corresponding components of the horizontal vector of current velocity; T_x and T_y are the components of the tangential stress exerted by the wind on the surface of the ocean. Furthermore,

$$\gamma_x = -\frac{\partial \zeta}{\partial x}; \quad \gamma_y = -\frac{\partial \zeta}{\partial y}; \quad b = kH - B; \quad B = \frac{gD}{4\pi\omega \sin\varphi}; \quad k = \frac{g}{2\omega \sin\varphi};$$

$c = \frac{1}{2\omega \sin\varphi}$, where ζ designates the elevation of the ocean's free surface above a certain horizontal plane XOY, parallel to the ocean bed and to the undisturbed surface of the ocean; g is the acceleration of gravity; ω is the angular velocity of the Earth's rotation; φ is the latitude; D is the so-called «friction depth» which, in turn, is connected with the coefficient of internal turbulent friction $A_z (A_z = \text{const})$ by the relation $D = \pi \sqrt{A_z / \omega \sin\varphi}$.

As we are concerned here with trade winds, we shall consider the velocity of the wind to be directed along the X-axis, and to vary only in the transverse direction Y. Seeing that L, the length of our contour as measured in the zonal direction, which is also the direction of the wind, considerably exceeds its breadth l, we may consider with sufficient accuracy that the elevation of the ocean's surface in the direction of the wind is a linear function solely of coordinate x, or, in other words, that γ_x is a constant quantity. As for γ_y let us agree that this quantity is a function of y alone. This implies that the motion under examination is the same in any cross-section of the considered region, transversal to the wind, which indeed approximates the conditions encountered in the region of the equatorial counter-current in the Pacific.

Under these assumptions expressions (1) may be rewritten as follows

$$S_x = B\gamma_x + b\gamma_y(y); \quad S_y = B\gamma_y(y) - b\gamma_x - cT_x(y) \tag{2}$$

Obviously, the following conditions must hold true for the case of steady motion

$$\int_0^l S_x\, dy = \int_0^l \int_0^H a\, dy\, dz = 0; \quad \int_0^L S_y\, dx = \int_0^L \int_0^H v\, dx\, dz = 0 \tag{3}$$

In so far as in the absence of inflow or outflow the initial volume of water in the considered region must remain constant, the, following condition must also be observed, in addition to conditions (3)

$$\int_0^l \zeta(y)\, dy = 0 \tag{4}$$

Using (2), (3) and (4) we can obtain an expression for the inclination of the ocean's surface in the direction of the wind

$$\gamma_x = \frac{c}{b}\left[\frac{\bar{T}B^2}{B^2+b^2} - \frac{2}{l^2}\int_0^l F(y)\, dy\right] \tag{5}$$

and also a profile of the ocean's surface in the direction transversal to the wind

$$\zeta(y) = \frac{cBl\bar{T}}{2(B^2+b^2)}\left(1-\frac{2y}{l}\right) + \frac{c}{B}\left[\frac{2}{l}\overline{F(y)}\,y - F(y)\right] \tag{6}$$

where $F(y) = \int T_x(y)\, dy$; $\overline{F(y)} = \frac{1}{l}\int_0^l F(y)\, dy$, and $\bar{T}$ is the mean value of T for a given region.

It follows from formulae (6) that the profile of the ocean's surface across the direction of the wind is the result of the superposition of two functions, characterized by two terms in the right-hand part of (6). The first, a linear term, represents the change in the level due to $\bar{T}$. The second term characterizes the effect of the transverse non-uniformity of the wind. We shall call it, therefore, «perturbation function» $\Psi(y)$.

To express the integrals in elementary functions let us assume that the tangential stress exerted by the wind in the region under consideration varies in accordance with the following simple law

$$T_x(y) = -\frac{|T_{\max}|}{2}\left(1+\cos\frac{2\pi y}{l}\right) \tag{7}$$

Expression (7) describes, of course, only the main features of the non-uniformity of the trade winds in the equatorial region of the Pacific; the details, unfortunately, remain unknown.

Substituting (7) in (6) we obtain after some transformations

$$\zeta(y) = \frac{cl|T_{\max}|}{2B}\left[\frac{-1}{4\pi^2\left(\frac{H}{D}\right)^2 - 4\pi\left(\frac{H}{D}\right) + 2}\left(\frac{1}{2}-\frac{y}{l}\right) + \frac{1}{2\pi}\sin\frac{2\pi y}{l}\right] \tag{8}$$

The velocity of the geostrophical current U_g in the direction of X is determined by the following expression

$$U_g = -k\frac{\partial \zeta(y)}{\partial y} = -\frac{kc\,T_{\max}|}{2B}\left[\frac{0.5}{2\pi^2\left(\frac{H}{D}\right)^2 - 2\pi\left(\frac{H}{D}\right) + 1} + \cos\frac{2\pi y}{l}\right] \tag{9}$$

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Since, according to (7), the velocity of the wind at a point $y = l/2$ is equal to zero, the actual velocity of the surface current at that point must be equal to the velocity of the geostrophycal current calculated from (9). It follows from (8), however, that the greater the initial ordinate of the linear change in the surface level compared to the amplitude $1/2\pi$ of the perturbation function, the smaller the influence of the latter, and even in the calm belt $(y = l/2)$ there may be no counter-current in some cases. In the Pacific, however, this possibility needs not be considered.

In examining the region of the equatorial counter-current in the Pacific, let us assume $\varphi = 5°$ N, and, following Defant (1), put $A_z = 10^2$ cm²/sec. Then it will be found that $D \cong 123$ m. Even if $H = 400$ m (the depth to which the equatorial counter-current system extends), not to speak of the actually much greater depth of the Pacific, $H/D \cong 3.3$, and the first term in (9) will have a value of 0.01, which (for $y = l/2$) can surely be neglected in comparison with unity. Therefore the maximum velocity of the surface counter-current may be determined with sufficient accuracy by the following expression

$$U_g = \frac{kc \, |T_{\max}|}{2B} = \frac{\pi \, |T_{\max}|}{2D\omega \sin \varphi} \tag{10}$$

In the Pacific the velocity of the trade winds is probably equal to 10 m/sec. This corresponds to a value of $T_{\max} = 3.2$ dynes/cm².

Substituting these values in (10) we obtain

$$U_g = \frac{3 \cdot 3.2}{123 \cdot 10^2 \cdot 128 \cdot 10^{-7}} = \frac{9.6}{15744 \cdot 10^{-5}} = 60 \text{ cm/sec.}$$

The value of U_g calculated in this way coincides quite satisfactorily with the results of observations, according to which the velocity of the equatorial current is little above 50 cm/sec.

It should be kept in mind that the actual velocity of the surface current U_0 is made up of the velocity of the geostrophycal current U_g and the velocity of the wind drift current on the surface $U_{0, w}$. The latter, according to Eckman, may be expressed as follows

$$U_{0, w} = \frac{\pi T'}{2D\omega \sin \varphi}$$

Hence, taking into account (7), we find

$$U_0 = U_{0, w} + U_g = \frac{-\pi \, |T_{\max}|}{4D\omega \sin \varphi} \left(1 + 3 \cos \frac{2\pi y}{l} \right)$$

whence it follows that the actual boundaries of the counter-current on the ocean's surface do not coincide with the transverse deformation of its surface, i. e. the deformation that determines the characteristic velocities of the geostrophycal counter-current U_g. To obtain the actual boundaries of the surface counter-current we must assume $U_0 = 0$ in the last expression. Then we shall find that $y_1 \cong 0.3 \ l$; $y_2 \cong 0.7 \ l$.

The upper part of the adjoined figure represents, in arbitrary units, the variation of T according to relationship (7), whilst the middle part of the same figure brings out the variation of the perturbation function $\Psi(y)$ which in this case coincides (within the accuracy of the figure) with the transverse profile of the ocean's surface $\zeta(y)$. As we see, this profile shows all of the main features typical of the surface variation in the Pacific in a cross-section transversal to the equatorial counter-current between the equator and 20°N(2). The broken line in the figure shows the linear variation of the level, which has practically no effect on the curve $\zeta(y)$. A and B denote the region of the actual surface counter-current which is directed against the wind at the edges. Finally, the figure shows the vertical boundaries of the counter-current from the surface to depth D, obtained by means of a simple graphical construction which will be explained in a subsequent paper. The general out-

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line of these boundaries coincides with the shape of the equatorial counter-current's vertical boundaries drawn up by Sverdrup ([2]) on the basis of hydrological observations.

Finally, it is interesting to ascertain whether or not the equatorial counter-current would exist in the absence of a calm belt between the trade winds.

With this object in mind let us assume that

$$T_x(y) = -\left(\frac{|T_{max}| + |T_{min}|}{2} + \frac{|T_{max}| - |T_{min}|}{2} \cos \frac{2\pi y}{l} \right) \qquad (11)$$

where T_{min} is the minimum value of the tangential stress of the wind in the centre of the considered region.

If it be taken into consideration that $T = \rho' x w^2$, where ρ' is the density of the air, x is a constant coefficient, and w is the wind velocity, the condition for the existence of a geostrophycal counter-current in the centre of the region $(y = l/2)$, and consequently of a deformation on the surface typical of this counter-current, may be written thus

$$\frac{w_{min}}{w} < \sqrt{ -\frac{0.5}{2\pi^2 \left(\frac{H}{D}\right)^2 - 2\pi \left(\frac{H}{D}\right) + 1} }$$
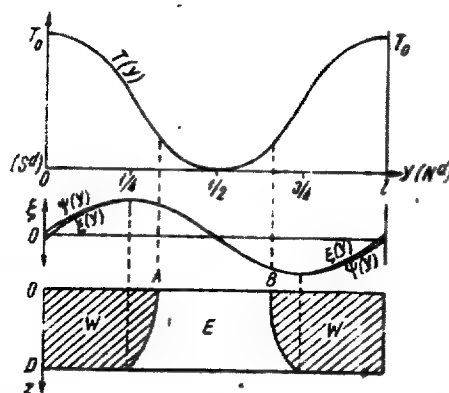
In the case where $H/D = 3.3$, the second term under the square root sign equals 0.01. Hence $w_{min}/w < 0.995$.

In other words, for the existence of a geostrophycal counter-current it is sufficient that the wind velocity in the centre of the region be only 1 per cent smaller than the average wind velocity!

It is not difficult to show that the possibility for the counter-current actually to exist on the surface under the same law of transverse non-uniformity of the wind (11) is determined by the criterion

$$\frac{w_{min}}{w} < \sqrt{\frac{2}{3}} \sqrt{ 1 - \frac{0.5}{2\pi^2 \left(\frac{H}{D}\right)^2 - 2\pi \left(\frac{H}{D}\right) + 1} }$$

If $H/D = 3.3$, $w_{min}/\overline{w} < 0.817$. From this one can readily understand why the equatorial counter-currents, caused — as we have shown — by the effect of the transverse non-uniformity of the wind, may in some places be found to move directly against the wind.

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ON THE PROMOTION OF CONTACTS BY MEANS OF METAL-ORGANIC COMPOUNDS

By G. M. ZHABROVA, S. Z. ROGINSKY, Corresponding Member of the Academy, and E. A. FOKINA

In course of investigations on the mechanism of promotion and the theory of preparation of active catalysts which are being carried out in the Catalysis Section of this Institute, it was found likely that catalysts might be able to be promoted by unstable volatile compounds. After a series of preliminary experiments we concentrated our attention on metal-organic compounds.

The investigation to be discussed here was carried out with tetraethyl lead. A copper-chromite contact was selected as the initial catalyst. It was prepared by the thermal decomposition of copper chromate, obtained by the interaction of copper nitrate and ammonium bichromate in ammonia solution. The black powder which resulted from the pyrolysis of the copper chromate was freed of excess copper oxide by treating several times with 10 per cent acetic acid, and was close in composition to copper chromite $(CuCr_2O_4)$, in accord with the data of Frazer [1] and others [2]. Examination with an electron microscope [3] revealed it to be a very friable and porous substance of distinctly crystalline nature.

For promotion the copper chromite was subjected to the action of chemically pure tetraethyl lead. The treatment was carried out in various ways, *viz.*, by soaking the catalyst in solutions of tetraethyl lead in hydrocarbons (pentane, heptane, octane) and then calcining it in air; by decomposing photochemically the tetraethyl lead soaked up by the catalyst from the hydrocarbon solution; by passing a stream of air containing vapours of tetraethyl lead over a heated and mobile layer of the catalyst. In the last method mentioned the operation was performed in a specially constructed glass apparatus with a porous partition fused into it, on which the catalyst was placed. By selecting a partition of definite porosity and varying the velocity of the air stream into the apparatus, the catalyst was kept mobile while passing the tetraethyl lead vapours through it. This method seems to afford a uniform distribution of the promoter between the grains of the catalyst.

Spectral analysis for the lead content in the samples of promoted catalysts, carried out in our laboratory by I.I. Levintov, showed that the catalyst captures about 1/8 of the metal let into the reaction apparatus in the form of tetraethyl lead.

The catalytic activity of the catalysts after promoting with tetraethyl lead was studied for three reactions: the decomposition of hydrogen peroxide, the oxidation of hydrogen and for the oxidation of iso-octane. The reaction with hydrogen peroxide was studied more elaborately than the other two. The kinetics of decomposition of the hydrogen peroxide was measured volumometrically in a shaking apparatus. The concentration of the aqueous solution of hydrogen peroxide usually employed in these experiments was 1.7 M, which corresponds to a content of 5.78 per cent by weight.

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The kinetics of oxidation of hydrogen was studied under static conditions in a vacuum apparatus, and the kinetics of oxidation of iso-octane, in a dynamic apparatus of the usual type.

As may be seen from Fig. 1, treatment with tetraethyl lead greatly increases the catalytic activity of the copper chromite. The catalytic activity of pure copper chromite in the decomposition of hydrogen peroxide at a temperature of 20°C is very low, but as the lead content rises the catalytic activity abruptly increases. This increase in activity continues up to a definite lead content in the catalyst; further increase of the lead content decreases the •
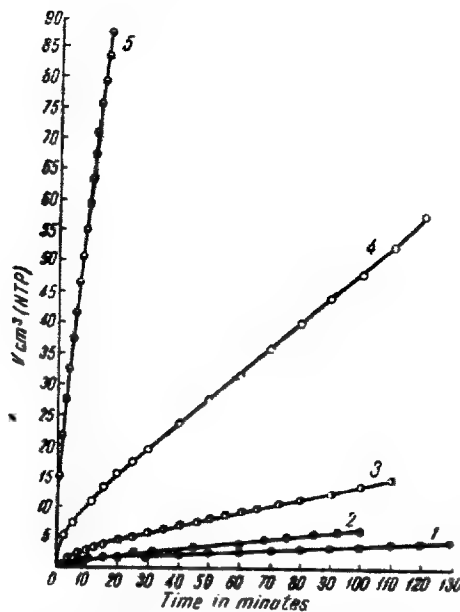
Fig. 1. Influence of treatment with tetraethyl lead on catalytic activity of copper chromite toward decomposition of hydrogen peroxide: *1*, pure copper chromite; *2*, copper chromite treated with Pb $(C_2H_5)_4$ by passing 0.001 g of the latter through a reaction vessel containing 1 g of catalyst; *3*, ditto with 0.007 g $Pb(C_2H_5)_4$; *4*, ditto with 0.045 g $Pb(C_2H_5)_4$; *5*, ditto with 0.090 g $Pb(C_2H_5)_4$
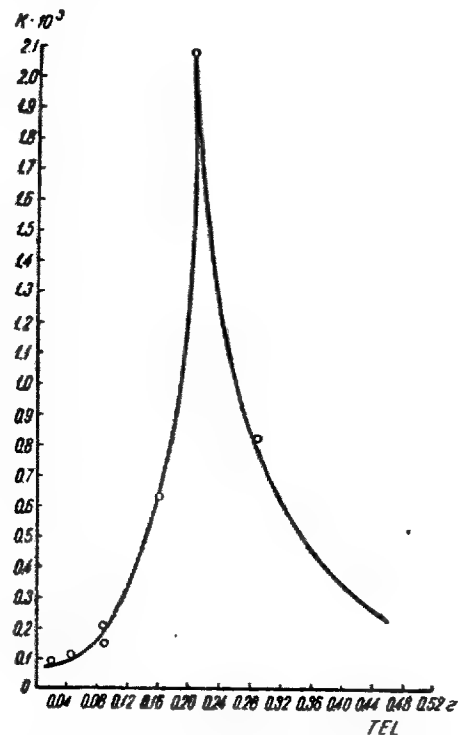
Fig. 2. Promotion curve of copper chromite when treated with tetraethyl lead with reference to decomposition of hydrogen peroxide.

activity of the catalyst. Lead alone and lead oxide have a poor catalytic activity.

A promotion curve of typical shape is shown in Fig. 2. The ascending branch of the curve may be said to obey an exponential law. The optimum promoter content in terms of lead oxide is about 3 per cent.

Further investigation of the influence of the promoter on the velocity of decomposition of hydrogen peroxide showed that introduction of the promoter causes an essential change in the kinetics of the reaction. Thus, the course of this reaction, which with pure copper chromite is linear at the outset and afterwards becomes parabolic (Fig. 3), changes when the promoter is introduced (Fig. 4). It may be shown that the velocity of decomposition of hydrogen peroxide over copper chromite promoted with lead conforms fairly well to an equation of the first order.

It should be noted that the activation energy of the decomposition of hydrogen peroxide ($\sim 20\,000$ cal/mole over pure copper chromite) changes comparatively little when the promoter is introduced.

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The kinetics of the oxidation of hydrogen over copper chromite was studied in a temperature range of 160 to 240° C. It was found that the course of the reaction fits in well with an equation of the first order. The catalyst is
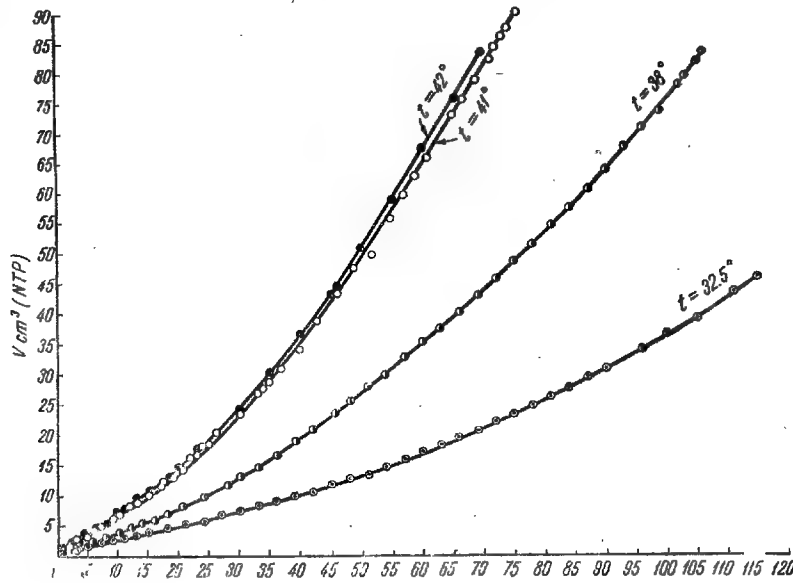
Fig. 3. Kinetic curves of decomposition of hydrogen peroxide over pure copper chromite at various temperatures. Conditions of experiment: 20 cm³ H_2O_2; catalyst, 0,3 g.

promoted when treated with tetraethyl lead (Fig. 5). The activity of the catalyst towards oxidation of hydrogen is at its maximum with a much smaller
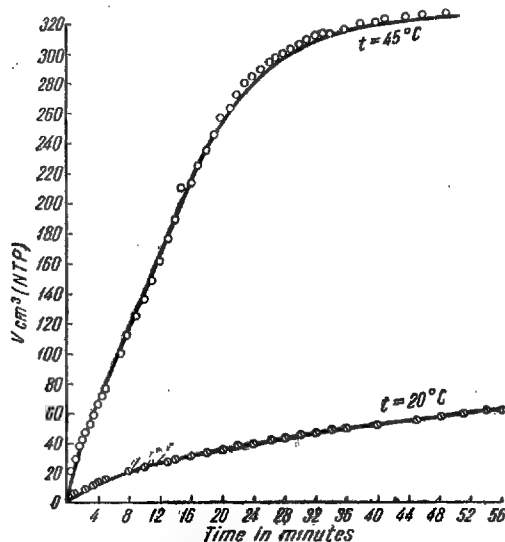
Fig. 4. Kinetic curves of decomposition of hydrogen peroxide over copper chromite after treating with tetraethyl lead. Conditions: 20 cm³ H_2O_2; catalyst, 0.03 g.
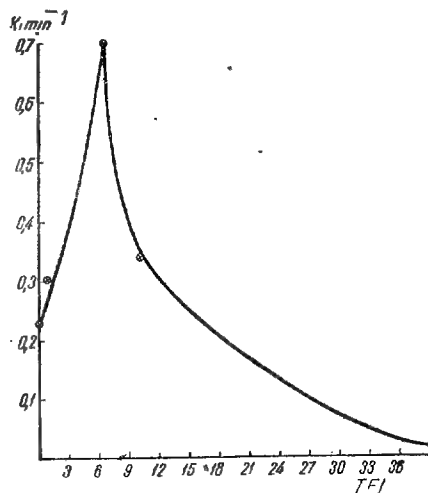
Fig. 5. Promotion curve of copper chromite with respect to oxidation of hydrogen after treating with tetraethyl lead.

promoter content as compared to its activity towards decomposition of hydrogen peroxide. The promoter content at the point of maximum activity, according to the data of spectrum analysis, is about 0.03—0.09 per cent in

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terms of lead oxide. Introduction of tetraethyl lead in quantities corresponding to the maximum activity of the catalyst for the decomposition of hydrogen peroxide has a deep poisoning effect on the reaction of hydrogen oxidation and results in deviation from the first order law.

It should also be noted that the maximum velocity of oxidation of hydrogen on the promotion curve is highly dependent on the method employed in treating the catalyst with tetraethyl lead.

The activation energy of the oxidation of hydrogen over copper chromite equals $17\,000 \pm 1000$ cal/mole. Introduction of the promoter does not change this value at the maximum of the promotion curve. The influence of tetraethyl lead on the oxidation of iso-octane is manifested both by an increase in the velocity of the reaction, and by a change in the pre-exponential factor and in the energy of activation. These coefficients may be found to increase or decrease, according to the method of introducing the promoter into the catalyst. A more detailed study of this phenomenon is under way.

It is early as yet to make assertions as to the possible range of application of this new method of promotion; we should only like to point out that tetraethyl lead is not alone among metal-organic compounds to give this effect. Elsewhere we shall discuss the possible mechanisms of the promoting action of the metal-organic molecules and the striking analogy between this new method of activation and our earlier observations ([4]) on the promotion of metallic films by occluded gas molecules.

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PHYSICAL CHEMISTRY

A NEW METHOD OF INVESTIGATION OF THE STABILITY
OF TWO-SIDED LIQUID FILMS

By A. M. SMIRNOVA and P. A. REHBINDER, Corresponding Member
of the Academy

The lifetime of liquid films before they burst and of single bubbles or drops in contact with each other or with an interface is a measure of the stability of the corresponding disperse system—emulsion or foam ([1]).

The lifetime of two-sided liquid films depends on the rate of decrease of film thickness due to the action of gravity and surface tension. The formation of adsorption layers of surface-active substances increases the life of films, because of the appearance of local two-dimensional pressures, hindering the decrease of film thickness, and also because of the direct improvement in mechanical (structural) properties of the film (viscosity, elastic-plastic properties—ultimate static shearing stress). The film usually bursts when its thickness falls to the critical minimum value, corresponding to the formation of the so-called «black spot».

The stability of bubbles on a water surface covered with a unimolecular layer of oleic acid, was first studied by W. Hardy ([2]) in 1925. P. A. Rehbinder and collaborators ([3]) and later on D. Talmud and S. Sukhowolskaja ([4]) used a similar method of measuring «elementary stability» in investigating the stabilizing action of surface-active substances in foams and emulsions.

Unfortunately, the poor reproducibility of separate values of the lifetime τ obtained by this method, makes it necessary to have a large number of separate measurements and to apply statistical methods. This is very troublesome and limits the application of the method. Besides, the observations become unreliable when made on films of short lifetime lasting but a few seconds.

In this connexion the authors have developed a method of evaluating the stability of films in «models» of emulsions or foams, by the maximum length of the film before rupture, when being drawn out of the liquid at a constant rate.

The device consists of two communicating cylindrical containers a and a' (see figure) 15 to 20 mm in diameter. One of them includes two platinum wires («rails») b running along its axis and welded into the bottom, which are connected by a cross-piece («stirrup») c. The upper parts of the containers are connected by means of a four-way cock d. To suppress the lack of reproducibility due to insufficient cleanness, the film is completely isolated from the outward medium (the glass tube is sealed up); an inert gas carefully purified from dust is used as the inside medium. It enters the containers a or a' through tube e and cock, and leaves it through tube f which lets it out into the atmosphere through trap l and rheometer k, controlling the rate of gas flow.

The device is filled with a liquid (or two liquids) through the ground (air-tight) connexion, and placed into a thermostat. The liquid film is formed between the platinum wires, when the interface liquid 1—liquid 2 or liquid—gas moving in the container a' upwards or downwards under the

3 C. R. Acad. Sci. URSS, 1946, v. LII, № 4.

pressure of the inert gas, passes the stirrup. Being caught by it, the film is gradually drawn out until it breaks, while the interface is displaced. The limit length (breaking length) of the film L_m varies with the velocity of displacement of the interface.

The lifetime of a film drawn out at a constant rate V may be estimated either directly, using a stop-watch for exact timing, or by measuring the maximum length of the film when breaking, and calculating τ from the relation $\tau = L_m/V$.

The method described has been successfully used in the investigation of emulsifiers and foam-producers, especially for low concentrations, $i. e.$ for systems having but a low stability. The reproducibility of separate measurements proves to be rather good, as the table shows.
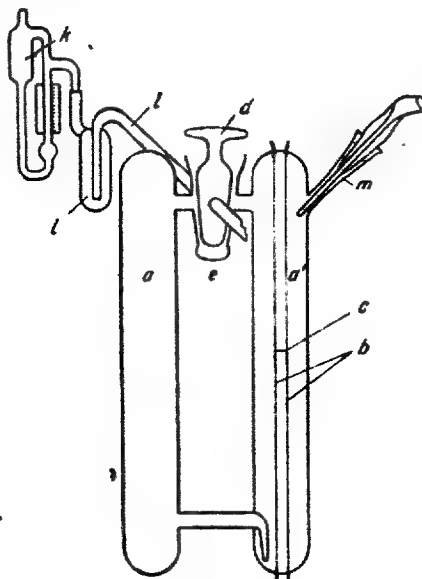
When studying a system of two liquids (water—oil) the stability of au emulsion of the «water-in-oil» type is estimated by drawing an oil film in the water medium (the liquid moves upwards). Contrary to this, the conditions of existence of a film in an oil medium (the liquid moves downwards) correspond to an emulsion of the «oil-in-water» type. By varying the type of emulsifier it is possible to judge of the transformation of phases in emulsions from the stability of water or oil films (*).

The table contains an example of the application of this method to the study of the stability of films formed in water from a vaseline oil with admixtures of oleic acid as a weak stabilizer. The velocity of extension of the film varied from 2.4 to 24 mm per sec.; within these limits the results were constant, $i. e.$ the ratio L_m/V was independent of the velocity.

Obviously, in such emulsion systems the effective thickness of the film h remains constant during the extension of the film up to the breaking point, and the process reduces in fact to an increase in the volume w of the liquid drawn out into a film of length L ($dw = hdL$) and of constant thickness. It should be noted, however, that the breaking of the film occurs, as a rule, in the vicinity of the stirrup.

In view of the invariability of the results obtained this method holds out promise of successful applications, especially in the study of emulsion systems of low stability. It is a great advantage, indeed, in such cases to be able to substitute the direct measurement of the lifetime with measurement of the maximum length h_m of the film at the instant of rupture when extended at a constant velocity. It is interesting to note that the method of observing

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Stretching in Water Medium of Films of Vaseline Oil
Added with Oleic Acid

Stretching velocity, V mm/sec.	Maximum length of film before rupture						Lifetime computed $\tau = L_m/V$
	separate values (successive tests)					mean	
0.01% oleic acid							
2.4	6	6	6	6	3	5.4	2.3
7.1	17	17	15	14	18	16	2.2
9.2	18	18	18	20	20	19	2.0
12.8	30	30	30	30	28	30	2.3
24.0	50	50	50	55	55	52	2.1
0.07% oleic acid							
6.0	28	25	25	26	25	26	4.3
8.8	38	38	36	40	37	38	4.3
11.3	45	46	48	48	47	47	4.15
15.5	63	65	70	65	63	65	4.2

the films under dynamic conditions is in itself to a certain degree an indication of the elementary act of formation and coalescence of drops in emulsions or of bubbles in foams.

When studying the stability of films formed by water solutions of alcohols in gas (vapour) media, the good reproducibility of results remained as before, but the invariability of the ratio L_m/V did not hold any more: the lifetime of the films decreased with an increase in the velocity of stretching. Obviously, in this case, the mechanism of formation and extension of the film was different and consisted in its thinning, especially at higher stretching velocities. In the limiting case the extension of the film over a certain critical length (or thickness) could be considered as a process of decreasing the film thickness at a constant volume of the film

$$dw = 0, \quad h = w_0/L$$

This limiting case is the opposite of «pure extension» of the liquid in the film. Clearly, in both cases the reasoning applies to the mean «effective» thickness of the film h. Actually, the coalescence process in emulsions as well as in foams consists in spontaneous l o c a l decrease in thickness and collapse of the film. When liquid films are formed by means of the device described in a gas medium, well pronounced interference fringes are observed in the films; these enable one to estimate the film thickness in different parts of the film area.

The lack of invariability for liquid films in a gas medium is not an obstacle to the application of the method. By means of it very interesting results are obtained, especially if the limit length of the film, stretched up to the breaking point, be measured as a function of the extension velocity. This can be achieved without difficulty owing to the reproducibility of individual measurements, whereby repetition on a large scale is rendered superfluous.

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BIOCHEMISTRY

PROTEIN BOUND PHOSPHATE AS A PRODUCT OF ENZYMIC HYDROLYSIS OF ADENOSINETRIPHOSPHORIC ACID

By E. T. SZÖRÉNYI and O. P. CHEPINOGA

(Communicated by A. V. Palladin, Member of the Academy, 15. XII. 1945)

According to the up to date conception, the enzymic hydrolysis of adenosinetriphosphate (ATP), catalysed by adenosinetriphosphatase, is considered to be the direct chemical source of energy for muscular contraction, as well as for certain other kinds of cellular work. Inorganic phosphorus, entering into organic compounds by way of phosphorolysis or in coupling with certain oxidation-reductions, is again liberated from these compounds through ATP, and in this process a certain amount of energy, originally accumulated at the formation of the same organic phosphate bonds is released.

It is generally assumed, though no experimental proofs of this opinion have been supplied, that this liberated phosphorus differs in no respect from the preformed inorganic phosphorus. It is true, Kalckar ([1]) has upon theoretical grounds postulated the existence of phosphomyosine, yet his idea was not substantiated experimentally, or further developed.

The purpose of the present investigation was to study the properties of inorganic phosphate, formed by enzymic splitting of ATP. The main subject of our research was myosine, which according to Lubimova and Engelhardt([2]) is not only inseparable from but even identical with adenosinetriphosphatase.

Myosine was prepared according to the technique used by Engelhardt and Lubimova, Szent-Györgi's ([3]) data as to the time of extraction being taken into account. It was twice precipitated with cooled distilled water. We used for our experiments a $m/10$ glycine—NaCl buffer pH $= 9.1$, proposed by Bailey([4]). Phosphorus was determined by Fiske and Subbarow's method; nitrogen, by Kjeldahl's micromethod.

E x p e r i m e n t   w i t h   I n o r g a n i c   P h o s p h a t e . Into a cellophane bag of 7—10 cm³ capacity a myosine solution was placed together with half a volume of $m/50$ K_2HPO_4. The external liquid consisted of a corresponding salt solution without myosine. When after 5—6 hours the equilibrium was reached, the external solution, containing no myosine, showed a higher content in inorganic phosphorus, as compared with the internal solution (Table 1). This was obviously a result of the well-known Donnan equilibrium.

The described experiment shows that inorganic phosphate added to a myosine solution is indeed freely diffusible and its distribution follows Donnan's law. As we found, myosine itself has no effect on the rate of phosphate diffusion.

The ultrafiltrate of the above-mentioned mixture contains about the same amount of inorganic phosphorus as the initial mixture. (The collodion filter for ultrafiltration was prepared by dissolving air-dried commercial collodion in a mixture of equal parts of alcohol and ether).

E x p e r i m e n t   w i t h   ATP. This was performed under identical conditions as described for the inorganic phosphorus, the only difference being that, instead of $m/50$ K_2HPO_4, a solution of sodium—ATP has been added, the quantity of which was calculated to make the final concentration of terminal phosphorus groups approximately equal to that of inorganic phosphorus in the preceding experiments. In this case, as well as in the preceding one, the visible equilibrium is established within 5—6 hours. However, considering that we dealt with two simultaneous processes: enzymic splitting off the terminal phosphorus and its diffusion, we prolonged the time of the experiment to 24 hours.

T a b l e   1

C o m p e n s a t i v e   D i a l y s i s   o f   I n o r g a n i c   P h o s p h a t e   (K_2HPO_4)   i n   t h e   P r e s e n c e   o f   M y o s i n e A_2

P content in γ per 1 cm³ of test after 24-hour compensative dialysis		Notes
in the cellophan bag	in the external vessel	
135	143	K_2HPO_4 added in the cellophan bag only
125	125	K_2HPO_4 in the external vessel only
283	283	K_2HPO_4 added inside and outside
152	183	Do
173	180	»
160	180	»
152 *	184 *	K_2HPO_4 added in the cellophan bag only
144	176	Do
152 *	184 *	K_2HPO_4 added in the external vessel only
152	176	Do

* 6-hour compensative dialysis.

After the equilibrium had been established, the internal myosine containing solution showed a larger amount of inorganic phosphorus, as compared with the external solution containing no myosine.

These experiments prove that the phosphate split off by adenosinetriphosphatase is not in a free state, but is partly bound to myosine, and consequently the bound part does not participate in the osmotic equilibrium.

Another way to detect the myosine bound phosphorus is given by ultrafiltration. When ATP was added to the myosine solution at pH = 9.1 (glycine—NaCl buffer), and after an incubation period, during which about 50 per cent of labile phosphorus is being split off, our mixture was ultrafiltrated, the greater part of the split off phosphorus was found to be retained and the ultrafiltrate contained much less phosphorus than the initial mixture.

By this method we investigated some properties of the myosine bound phosphorus. It was found that by acidifying with acetic acid the myosine is precipitated, releasing bound phosphorus which passes into the ultrafiltrate. The same effect is observed after the denaturation of myosine by heat. On the contrary, the phosphorus binding capacity of myosine is increased by alkalification (with several drops of concentrated alkali); attention should be paid to the fact that simultaneously preformed inorganic phosphorus is also bound. Evidently, enolysed protein is able to bind inorganic phosphorus.

For the quantitative determination of bound phosphorus we combine two methods: compensative dialysis and ultrafiltration. After 24 hours of compensative dialysis, the contents of the cellophane bag were divided into two parts.

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Compensative Dialysis of Phosphate Splitted off ATP in the Presence of Myosine

Added			P content in γ per 1 cm³ of test after 24-hour compensative dialysis					Bound P in mg per 1 g of protein	
in the cellophan bag		in the external vessel	in the cellophan bag		in the external vessel		bound	compensative dialysis method	ultrafiltration method
myosine	addition		P_0	P_{7-0}	P_0	P_{7-0}	P_a		
A_3	Ca ATP	—	185	127	89	119	96	—	—
A_3	Ca ATP	—	122	192	97	187	25	—	—
A_3	—	Ca ATP	82	162	66	170	16	—	—
A_1	—	Ca ATP	166	30	120	120	46	10.5	—
A_1	Ca ATP	—	210	40	163	30	47	10.5	—
A_1	—	Ca ATP	163	17	123	68	40	8.9	—
	Ca ATP	—	228	0	123	68	105	·24.4	—
A_3	Ca ATP	Ca ATP	280	87	97	103	183	—	—
B_1	Ca ATP	—	248	3	140	3	108	9.3	—
B_1	Ca ATP	Ca ATP	360	17	100	47	260	22	—
B_1	Na ATP	Na ATP	236	0	153	103	83	7.1	—
A_3 at pH=6.11	Ca ATP	Ca ATP	57	383	32	293	25	—	—
A_3*	Na ATP+CaCl₂ .	Na ATP	135	0	96	76	39	5.1	7.3
A_3*	Na ATP	Na ATP	124	4	108	56	16	2.1	10.4
A_3	Na ATP+CaCl₂ .	Na ATP	192	24	128	80	64	14.2	13.0
	Na ATP	Na ATP	192	16	156	60	36	8.0	13.3
A_3*	Na ATP+CaCl₂ .	Na ATP	104	32	64	116	40	15.8	—
A_3*	Na ATP+CaCl₂ .	—	100	56	76	28	24	11.4	—
A_3*	CaCl₂	Na ATP	54	10	30	162	24	11.4	—

One of them was subjected to ultrafiltration without any preliminary treatment (F_1), to the other 1—2 drops of glacial acetic acid were added before ultrafiltration (F_2). F_2—F_1 gives the value of the bound phosphorus, which within the limit of error coincides with the bound phosphorus value obtained by compensative dialysis.

The value of bound phosphorus at pH=9.1 equalizes 10—15 mg P per 1 g of myosine. This value does not depend upon myosine concentration. (For instance, a myosine solution containing 3.13 mg of protein nitrogen in 1 cm³ is able to bind 14 mg of phosphorus per 1 g of protein; after a twofold dilution of this solution 1 g of protein binds 15 mg of phosphorus).

The activation of adenosinetriphosphate by $CaCl_2$ produces but an insignificant effect upon phosphate binding. At pH=6.1 a qualitatively similar effect is observed. Myosine of the 1st precipitation, as well as myosine «B» of Szent-Györgyi (myosine + actine), both containing myokinase, do not differ from myosine of the 3rd precipitation in their effect upon phosphate binding.

Finally, an effect quite similar to that of myosine is observed also on the water-soluble adenosinetriphosphatase, prepared according to Kalckar's method [5] from water extract of muscle by means of isoelectric precipitation at pH=6.0.

The question of the character of the bond and the concomitant problem of the mechanical-chemical meaning of the described effect is still to be studied. For the time being, we can attribute to the bound phosphorus 2 physiological functions.

1) Bound phosphorus may be considered as a latent source of osmotic

power, when taking no part in osmotic equilibrium, it can be easily released and produce osmotic pressure, and *vice versa*.

2) Maintenance of an unequal distribution of inorganic phosphorus between the cellular and intracellular spaces.

The intracellular concentration of inorganic phosphorus, as it is known, exceeds many times that of extracellular phosphorus. The transition of the latter into the cells has been studied by many authors during recent years[5] with the aid of radioactive phosphorus. Inasmuch as 1 g of myosine is able to bind 10—15 mg of phosphorus, which is at least 5 times more than sufficient for retaining the whole intracellular phosphorus in a bound form, it seems to us most probable that the sparingly diffusing intracellular phosphorus studied by the authors mentioned is bound to protein.

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CHEMICAL TECHNOLOGY

SETTING TIME AND STRENGTH OF α-HEMIHYDRATED GYPSUM AS AFFECTED BY ADDITION OF SHERESH (*EREMYRUS SPECTABILIS*)

By P. P. BUDNIKOV, Corresponding Member of the Academy

Hemihydrated gypsum (plaster of Paris) produced by burning the dihydrate, boiling in tanks and calcining in shaft or rotary kilns, always contains more or less of the initial dihydrate.

The same is true when the hemihydrate is produced by treating gypsum with steam under pressure and drying. It is scarcely possible under industrial conditions to produce a product which would correspond to the formula of the hemihydrate, $CaSO_4 \cdot 0.5\,H_2O$, and contain the theoretical amount of water, *viz.* 6.2 per cent. In most cases the finished product contains at least three or four modifications of calcium sulphate, *viz.* the α-hemihydrate, the β-hemihydrate, the anhydrite and the dihydrate.

The presence of the dihydrate in the finished product considerably accelerates its setting time. This is due to the fact that the dihydrate particles act as crystallization centres, and as such, make the $CaSO_4 \cdot 2H_2O$ crystals separate out more rapidly from the supersaturated solution of the hemihydrate. In the case of a hemihydrated gypsum, especially $\alpha = CaSO_4 \cdot 0.5\,H_2O$, produced by treating the dihydrate with humid steam under pressure, and having as small a content of $CaSO_2 \cdot 2H_4O$ as 1—2 per cent, the setting process begins after 2—3 mins. and is completed in 3—5 mins.

Table 1

Setting Time and Strength of Plaster of Paris as Related to its Dihydrate Content

	Setting time, mins.		Ultimate compression strength, kg/cm²	
	beginning	end	after 1 day	after 7 days
100% hemihydrated gypsum	18	27	5.5	14.2
99% » » + 1% $CaSO_4 \cdot 2H_2O$	5	14	6.5	12.5
97% » » + 3% $CaSO_4 \cdot 2H_2O$	2.5	5	6.7	10.2
95% » » + 5% $CaSO_4 \cdot 2H_2O$	2.5	4.5	5.4	11.5
90% » » +10% $CaSO_4 \cdot 2H_2O$	2	4	3.4	8.7

Table 1 shows that as the content of $CaSO_4 \cdot 2H_2O$ in the hemihydrated gypsum rises, the setting time considerably decreases. The hemihydrated gypsum was prepared by grinding and burning natural transparent gypsum of the Kamenetz-Podolsk deposit at 150°C for a period of 2 hours, the dihydrate water content after burning being 5.7 per cent. The same natural transparent pure gypsum was used as an admixture. The fineness of the grind was ~9.8 per cent on 900 op/cm².

Plaster of Paris that has a very short setting time is unsuitable for manufacture of building parts or for other purposes. The setting time can be protracted by increasing the water-plaster ratio, but this considerably weakens the plaster, as can be seen from Table 2.

Table 2

Setting Time and Strength of Plaster of Paris as Affected by Water-plaster Ratio

W/P ratio	Setting time in mins. of the hemihydrate, prepared by boiling, under a pressure of 1.3 atm.		Ultimate compression strength, in kg/cm²
	beginning	end	
0.40	3.5	6.5	258
0.45	4.5	7	220
0.50	6.5	8.5	213
0.55	7	9	182
0.60	7.5	10.5	158
0.65	8.5	11	136

Table 2 shows that as the water : plaster ratio increases, the setting interval is protracted, but at the same time the strength of the plaster upon solidification decreases.

The most active mineral decelerator of the setting process in gypsum is borax. An admixture of only 0.2—0.3 per cent borax is sufficient to cause a considerable delay in the setting time. The decelerating influence of borax is due to an exchange reaction which takes place on the surface of the plaster particles:

$$CaSO_4 \cdot 0.5H_2O + Na_2B_4O_7 \cdot 10H_2O + aq = CaB_4O_7 + Na_2SO_4 \cdot 10H_2O + aq$$
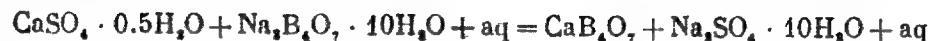
The grains of the hemihydrated plaster become coated with CaB_4O_7, and this, together with the decrease in the solubility of the plaster in the presence of sodium sulphate delays the setting time [1].

In building practice, however, borax is rarely used as a decelerator, as it is too expensive. A keratinous powder is used instead, which is made by treating horns and hoofs of animals with a hot sodium hydroxide solution and then adding ground quicklime.

In some cases carpenter's glue is used as a decelerator, but this is not always economical. There are indications in the literature [2] as to the possibility of adding very small quantities of sheresh to «ganch» [3] with the result that not only is the setting time decreased and the strength of the solidified parts increased, but the water resistance of the latter becomes higher as well. In our studies we have investigated the influence of sheresh on the setting time of high quality plaster of Paris, which ordinarily begins setting after 2—3 mins, and is set in 4—6 mins, making it very difficult to use in the manufacture of building parts. At the same time we studied the effect of this admixture on the strength of the articles made of such plaster.

Sheresh was kindly put at our disposal by A. S. Vatzeko (who got it from Baku). Sheresh (called shreshi in Georgian; shresh, in Armenian; cheresh or djirshi, in Tatar) grows in the Transcaucasus. It is very common in Azerbaijan, Armenia, Georgia and all over Middle Asia. It is a plant of the *Liliaceae* family, genus *Eremyrus* (*Eremyrus spectabilis*)[4]. The roots of this plant are dried and crushed to powder. A photomicrograph showing the microstructure of pulverized sheresh mixed with water is given in the figure (magnification, 80).

When mixed with hot water sheresh forms a good glue (with a slightly acid reaction), which is used in the Caucasus and other places by cobblers,

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for pasting wallpaper and also in bookbinding. The glue contains 27—33 per cent calcium arabiate. The composition of sheresh ash is as follows: SiO_2, 31.11 per cent; CaO, 24.83; MgO, 4.49; SO_2, 2.54; Cl, 5.11, *etc.* *
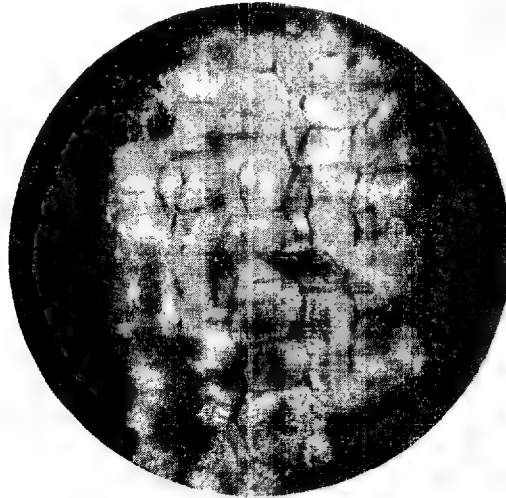
Table 3 illustrates the effect of adding sheresh to α-hemihydrated gypsum prepared by boiling pulverized natural gypsum of the Ust-Kamsk deposit under a pressure of 1.3 atm., and to α-hemihydrated gypsum prepared by treating gypsum from the Voroshilovgrad deposit with humid steam under a pressure of 1.3 atm. In the first case the water: plaster ratio was 0.45, and in the second 0.35.

Table 3

Sheresh added, %	Setting time, mins		Ultimate compression strength, kg/cm² *	Sheresh added, %	Setting time, mins		Ultimate compression strength, kg/cm² *
	begin-ning	end			begin-ning	end	
α-hemihydrated gypsum prepared by boiling dihydrous gypsum from the Ust-Kamsk deposit under pressure (W/P 0.45)				α-hemihydrated gypsum prepared by treating gypsum with steam under pressure in an autoclave apparatus (W/P=0.35)			
—	4.5	5.5	220	—	4	5	300
0.1	4.5	6	225	0.1	5	7	335
0.25	10	14	245	0.25	11	15	345
0.5	13.5	17.5	320	0.5	15.5	18	383
1.0	20	24.5	273	1.0	20	26	360

* The ultimate strength was determined as follows: the samples were taken out of their moulds 1 hour after filling, a day later the samples were dried in a drying oven at 50°C to constant weight.

These data show that a 0.2 per cent admixture of sheresh to α-hemihydrated gypsum is sufficient to cause an appreciable delay in the setting time. Increasing the quantity of sheresh added up to 0.5 per cent not only increases the delay in the setting time still further, but improves the strength of the solidified plaster as well. Further increase of the quantity of sheresh up to 1 per cent delays the setting time still more, but decreases the strength of the plaster.

* These data were received by the author from S. M. Weller (Tbilisi).

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We also studied the influence of sheresh on the plaster when boiled under a pressure of 1.3 atm. The data of this study are given in Table 4.

Table 4

Sheresh added, %	W/P ratio	Setting time, mins		Ultimate compression strength, kg/cm²
		beginning	end	
0	0.45	4.5	5.5	220
0.2	0.45	28	31	357
0.5	0.45	85	115	290

If the water: plaster ratio is decreased, the addition of sheresh to α-hemihydrated gypsum causes a considerable increase in its strength, as may be seen from Table 5.

Table 5

Sheresh added, %	W/P ratio	Setting time, mins		Ultimate compression strength, kg/cm²
		beginning	end	
0	0.47	2.5	4	191
1.0	0.33	22	29	456

Thus, sheresh is a valuable admixture for hemihydrated gypsum, not only delaying the setting time of the latter, but increasing its strength as well. A point of special interest is the introduction of sheresh while the plaster is being prepared by boiling pulverized gypsum under a pressure of 1.3 atm. This results in a considerable delay in the setting time and greatly increases the strength of the solidified plaster.

The influence of sheresh as a decelerator is probably due to the fact that the swelled sheresh prevents the formation of crystalline structure in the dihydrated gypsum while it is in the colloidal state. The increase in the strength of plaster castings when sheresh is added is apparently due to the dihydrated gypsum forming larger crystals and to the adhesive properties of the sheresh.

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GEOLOGY

ON THE PROBLEM OF THE GENESIS OF JAROSITES

By B. K. BRESHENKOV

(Communicated by D. S. Beliankin, Member of the Academy, 22. III. 1946)

Sodium, potassium and soda-potash jarosites $(Na,K)_2O \cdot 3Fe_2O_3 \cdot 4SO_3 \cdot 6H_2O$ are rather widespread in nature. In the USSR they occur in the zones of oxidation of many ore deposits. In the southern termination of the Ural Range, in North and Central Kazakhstan, in the Ore-bearing Altai jarosites are the typical minerals of the oxidation zone of nearly all the sulphide deposits ([5],[10]). In those regions, along with iron hydroxides, they are the principal iron minerals of the weathering crust.

Particularly large accumulations of jarosites in the form of solid masses of thousands of tons occur in the oxidation zones of certain large pyrite deposits of the South Urals (Blyava) and North-Eastern Kazakhstan (Maikain, Djusaly). Such accumulations sometimes carry a higher tenor of gold (Maikain, Djusaly), being commercial gold ore ([6]). Besides, solid orbicular and loaf-shaped concretions of jarosite up to 1 m in diameter from some of the deposits, *e. g.* the Maikain, may be used as excellent raw material to produce, by grinding and calcination, crocus, a fine polishing material (K. A. Nenadkievich).

Less widespread in the oxidation zones of the sulphide deposits are plumbojarosites $PbO \cdot 3Fe_2O_3 \cdot 4SO_3 \cdot 6H_2O$ ([5],[8],[10]), argentojarosites $Ag_2O \cdot 3Fe_2O_3 \cdot 4SO_3 \cdot 6H_2O$([8]). The occurrence of selenojarosites is possible; sometimes the selenium content in the total mass of the jarosites may be somewhat higher than in the other products of oxidation of sulphide ores ([5],[10]).

Jarosites are known to form: 1) in the process of oxidation of sulphides in the zone of hypergenesis, and 2) under conditions of solfataras. The formation of jarosites in the zone of hypergenesis is usually regarded as an intermediate phase of the process of oxidation of pyrites when they pass into iron hydroxides ([1],[8],[5],[8],[10−14]). In its simplest form the scheme of this process is as follows. Pyrite is oxidized to ferrous sulphate $FeSO_4$, which in the presence of free oxygen easily passes into ferrisulphate $Fe_2(SO_4)_3$. The latter is hydrolysed and through basic sulphates (jarosites) passes into iron hydroxides ([1]). Certain writers consider this process to be more complex; nevertheless, the formation of jarosites is in all cases considered as one of the phases in the process of oxidation of pyrite.

As a result of the study of the composition and structure of the oxidized ores of the pyrite deposits of Kazakhstan, the present author has arrived at the conclusion that the formation of jarosites may possibly have proceeded also in a different way, namely, that jarosites may have been regenerated at the expense of the iron hydroxides of the ancient «iron hats» under conditions of hypergenesis changed in the course of time: upon the transgression of the sea or of a considerable rise of the base level of erosion and of the level of underground waters.

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Such a mode of formation of jarosites is made possible by the reversibility of the well-known reaction

$$Fe_2(SO_4)_3 + mH_2O \rightleftarrows 2Fe(OH)_3 + 3H_2SO_4 + nH_2O \qquad (A)$$
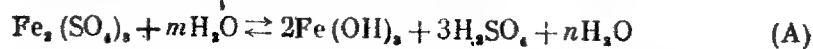
while its probability is proved by the geological conditions of the occurrence of oxidized ores, the Meso-Cenozoic geological history of the areas of development of jarosites, the composition of underground waters in deposits with large accumulations of jarosites and the textural peculiarities of the latter.

In the South Urals, North Kazakhstan, Ore-bearing Altai and in other regions the lower boundary of oxidation zones in many ore deposits is situated much lower than the present underground water table ([2],[4],[8],[10]). The difference of marks attains 20—35 m and more (Blyava, up to 25 m; Maikain, up to 30 m). This phenomenon, viewed in the light of the geological history and palaeogeography of the above-named regions during the Mesozoic, is accounted for by a deeper position of the ancient table of underground waters([10]).
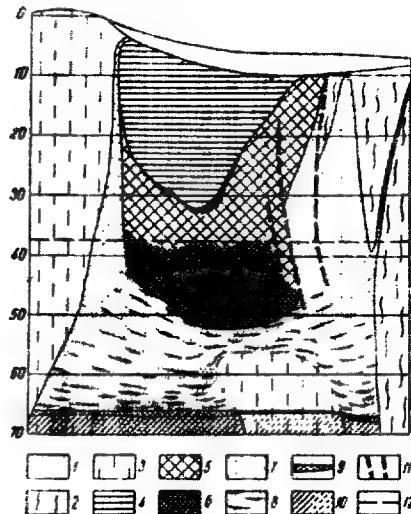
Fig. 1. Cross-section of the oxidation zone of the «C» ore-body of Maikain (workings not shown): 1, quaternary loams; 2, quartz-sericite schists; 3, «secondary quartzites»—albitophyres completely transformed into quartz; 4, variegated clays of the sagging funnels; 5, limonites of the «iron hat»; 6, jarosites; 7, quartz baryte ores; 8, quartz-baryte «quicksands»; 9, pyrite «quicksand»; 10, sulphide pyritic ores, nearly solid and disseminated; 11, recent tectonic disturbances; 12, underground water table.

In the present zones of oxidation of the pyrite deposits considerable masses of jarosites are situated below the level of underground waters(Fig.1). Higher up they are found to be replaced by limonites of the «iron hats».

During the Mesozoic, with a lower position of the underground water level and a prolonged period of chemical «weathering», in the zones of actual jarosite occurrence must have existed conditions necessary for a more complete oxidation. The access of free oxygen must have caused the above (A) reaction to proceed from left to right, i.e. to bring about, in the course of time, the transition of ferrisulphates and jarosites to iron hydroxides, similar to what is observed in the present oxidation zones above the level of underground waters.

It may therefore be presumed that the iron hydroxides in the zones of oxidation of many sulphide deposits in the regions named above, extended to considerably greater depths than they do at present.

In the South Urals and in North Kazakhstan during the Upper Cretaceous epoch and in particular in the Oligocene a transgression of the sea ([1],[9]) must have sharply changed the course of the processes of hypergene alterations.

At the period preceding the transgression the rise of the level of underground waters distinguished in pyrite deposits even under the actual conditions by the presence of free sulphuric acid and not infrequently $pH < 3$, must have caused the (A) reaction to proceed in a reverse direction—from right to left, i.e. towards a transition of iron hydroxides to colloidal solutions of ferrisulphate.

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This stage in the history of the oxidation zones of pyrite deposits having an insignificant content of vein minerals in the primary ores, is highly important, for it enables us to ascertain the cause of the sharp differentiation of the products of their oxidation to the modern «iron hats» and the «quicksands» underlying them.

With a transition of the iron hydroxides of the ancient zones of oxidation to colloidal solutions, highly favourable conditions were created for a gravity separation of the liquid phase in the form of colloidal solutions, on the one hand, and the solid phase in the shape of disintegrated particles of quartz, baryte, as well as gold particles and other scarcely soluble constituents, originally disseminated in the mass of pyrites and not combined into a solid skeleton,—on the other. These particles, precipitating to lower horizons of the
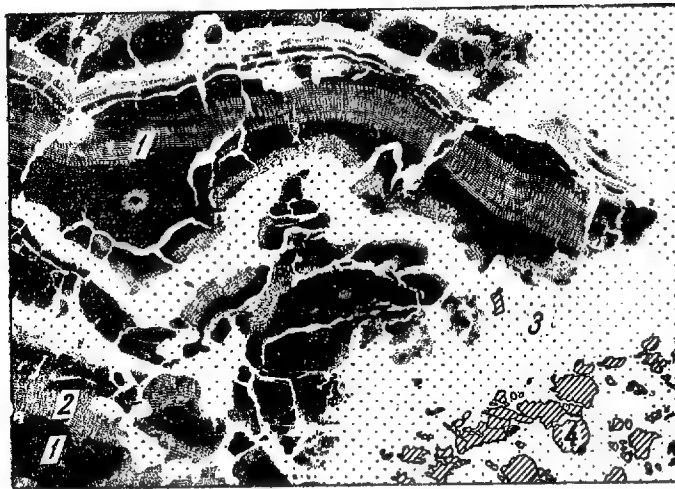
Fig. 2. Texture of jarosites with inclusions of limonites. Periphery of the «C» ore-body of Maikain, depth 30 m. Polished section; ×18. *1*, limonites of various composition; *2*, partly jarositized limonites; *3*, jarosites; *4*, grains of baryte and quartz.

oxidation zones produced accumulations of quartz-baryte «quicksands», or, as they are called in the Urals, «sands» enriched in gold.

Such an origin of the «quicksands» is evidenced by their horizontal bedding. The same is testified by the high concentration of shapeless gold particles, silvery inside and with a high value from the outside, on lenses of consolidated «quicksands» but not under them (Breshenkov, in press).

At the same time the colloidal solutions were cleared from mechanical admixtures, and it is only in this way that a most natural explanation may be given to the colloform texture and high purity of the loaves of jarosites, *e. g.* in the central part of the «C» ore-body of the Maikain deposit, which do not contain even minute inclusions of scratching quartz, or baryte particles.

Further . on, upon the transgression of the sea and mixing of acid underground mine ($pH < 3$) with salty sea waters ($pH = 7—8$), owing to a decrease in the acidity of the former there must have occurred coagulation of sols and additional adsorption by gels of the ferrisulphate of alkaline metals both from the wallrocks and from the sea waters.

It is with this stage that—as it would seem—the formation of silver halides AgCl, Ag (Cl, Br), AgI, *etc.* should be associated; these have been detected in many deposits containing jarosites and are there the principal form of silver compounds in the oxidation zones ([6],[10]). Chlorine, bromine and iodine were possibly also the main solvents of dispersed gold and in this way favoured

its redeposition in the upper horizons of the oxidation zones in the form of crystals and high grade films. In the jarosites and limonites of such deposits there occurs high-grade «secondary» gold redeposited from solutions, primarily in the cavities between individual concretions—loaves ([6]).

At last, of decisive importance for ascertaining the genesis of jarosite accumulations were the results of the study of the texture of compact varieties of the jarosites of the Maikain deposit carried out by the author in 1945.

In the «C» ore-body, at a depth of 30 and 50 m from the surface, *i. e.* 7 m above and 13 m below the level of underground waters, before the beginning of exploitation, in the peripheral parts of the ore-body, among quartz-baryte rocks, within areas of rather compact jarosites, were discovered acute angular and rounded fragments of compact varieties of «limonites» of a various composition, from hydrohematite to stilpnosiderite. These fragments proved to be thin stringers of jarosite broken along cracks (Fig. 2). A partial jarositization of various intensity of some fragments was also observed; it is outwardly pronounced in a reddish and yellowish coloration acquired by these fragments.

The textural peculiarities of the jarosites, the relics preserved in them are sufficient evidence of the formation of the jarosites of the «C» ore-body of Maikain at the expense of iron hydroxides, probably in the above-described fashion. It is but natural to suppose that such regeneration jarosites are in general common under the complex conditions of oxidation of pyrites.

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GEOLOGY

ON THE PRESENCE OF MARINE FAUNA IN A GYPSUM BAND

By O. S. VIALOV

(Communicated by V. A. Obruchev, Member of the Academy, 26. III. 1946)

The Turkestanian stage of the Palaeogene of Ferghana is made up by green gypsiferous clays divided into two parts (Trk_1 and Trk_3) by a horizon of limestones and marls (Trk_2) ([1]) exposed in the outcrops. The latter horizon is known in the literature also by the name of the «m» horizon, according to the nomenclature of K. P. Kalitzky. The whole formation is characterized by a peculiar and typical fauna, primarily of oysters, belonging mainly to the genus *Fatina*. This genus, founded at a comparatively recent date, comprises those representatives of the subfamily *Gryphaeinae*, the lower valves of which, as distinguished from *Gryphaea* proper, show a radial sculpture, however slightly pronounced ([2]). Among the characteristic forms of the Turkestanian stage belong *Fatina esterházyi* Pav., *F. böhmi* Vial. and their varieties, *Liostrea (Kokanostrea) kokanensis* Sok., and others.

In the upper member of green clays (Trk_3), which are highly gypsiferous, within 1 m from the base, lies a rather persistent horizon, 5—10 cm thick, consisting of large-sized and most typical *Fatina esterházyi* Pav. and carrying likewise numerous *Ostrea (Flemingostrea) schurabica* Vial. and *Chlamys vialovi* Korob. The oysters are generally highly gypsified, and the horizon itself is distinguished in the exposures as a band of gypsum several centimetres thick, stained brown by iron hydroxides.

The joint occurrence of oysters and pecten in a gypsum horizon appeared rather strange, as they are known usually to inhabit sea water of normal salinity.

In 1943, when visiting the Maili-Sai petroleum and ozokerite deposit the author had the opportunity to examine freshly sunk test-pits which exposed the lower part of the Trk_3 green clays to the underlying Trk_2 horizon of limestones («m» horizon of K. P. Kalitzky). The above-mentioned horizon with large-sized *Fatina esterházyi* Pav. was also discovered here.

It has been found in the first place that when this horizon does not outcrop on the surface but is opened up by a test-pit at a certain depth, gypsum is absent there and the oysters are not gypsified. The green clays are found to carry a large number of small pyrite crystals. Similar pyrite crystals cover the surfaces of the valves of the oysters obtained from the test-pit. Later on, V. N. Muratov brought to the author a similar pyritized *Fatina esterházyi* Pav. from a test-pit of the Shor-su ozokerite deposit (South Ferghana). In the test-pit, *i. e.* far from the surface, gypsum is also absent.

It is thus quite evident that gypsum is secondary in this horizon, being a product of pyrite decomposition, as a result of a joint work of oxygen and water in the presence of calcium carbonate. The process is quite common and universally known. The totality of the reactions taking place here; *viz.*

oxidation, hydration, interaction of the sulphuric acid formed with carbonates (calcite), may be reduced to the following general scheme (V. A. Uspensky):

$$FeS_2 + O_2 + H_2O + CuCO_3 \rightarrow CaSO_4 + Fe_2O_3nH_2O + CO_2$$
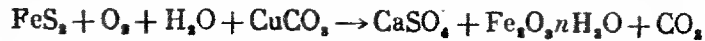
Thus, the decomposition which takes place at the surface of pyrite is accompanied by the formation of gypsum, on the one hand, and of brown aqueous iron oxides to which the latter owes its colour, on the other. If large amounts of pyrite are present, a loose interlayer of brown gypsum is formed on its surface. The same decomposition proceeding at the surface of pyrite enclosed in the whole bed of the green clays of Trk₂ is responsible for their being gypsiferous, this character being secondary.

The apparent ecological incongruence of a joint occurrence of oysters and gypsum is thus accounted for in a satisfactory way: the oysters have never inhabited the medium in which the deposition of gypsum took place. They lived in a normal marine medium, while the gypsum was formed as a result of the subsequent processes of pyrite decomposition.

From the aforesaid another conclusion is to be drawn. When analysing the conditions of formation of definite sediments, in particular, of clays, if devoid of fauna, as an argument in favour of marine origin the presence of gypsum has often been cited. Above it has been shown that such an occurrence of gypsum may be of a secondary nature, and not at all associated with marine conditions. This feature should therefore not be considered as evidence of the marine origin of the sediments in question.

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MINERALOGY

POLLUCITE IN PEGMATITES OF THE KALBIN RANGE (EASTERN KAZAKHSTAN)

By A. I. GINSBURG

While studying the mineralogy of lithium pegmatites of the Kalbin Range (Eastern Kazakhstan) the present author discovered in 1943 in the Ungursay spodumen deposit considerable accumulations of a white mineral bearing strong resemblance to quartz; upon further examination this mineral was found to be caesium alumosilicate, the so-called pollucite. This find evoke a great interest in industry, since pollucite proved a highly valuable raw material for obtaining caesium. Prospecting and exploration operations in this region have resulted in a discovery of a new pollucite occurrence (Krasno-Kordon) by M. J. Abdulina, geologist.

At present pollucite is known to occur in several veins of the Kungursay and Krasno-Kordon deposits. Both these deposits are located in the central portion of the Kalbin Range and are constituents of the Asu-Bulak pegmatite field traced along the valley of the Asu-Bulak, a right-hand tributary of the Tainta River (a left tributary of the Irtysh), and located 90 km SE of the city Ust-Kamenogorsk. Pegmatite veins of this region occur in medium-grained porphyritic biotite granites of the Variscian age.

Similar to all other rare earth minerals, pollucite in pegmatite veins is confined to the axial parts of large swells and occurs together with lepidolite, pink and polychromous tourmaline, amblygonite, spodumen or petalite, white alkaline beryl and a mineral of the microlith group. Pollucite forms no regular crystals and is encountered as large solid and granular accumulations of an irregular shape. The biggest excretions of pollucite discovered by M. J. Abdulina in the Krasno-Kordon deposit measure up to 1 m across.

Pollucite is white in colour, separate grains being transparent and colourless; on weathering pollucite in some instances turns yellow or cream-coloured (vein No. 3 of the Ungursay deposit). Pollucite excretions show mostly fine-grained structure, whereas solid massive varieties are rarely encountered (Krasny-Kordon). Separate small ice-transparent grains with shell-like fracture are frequently observed under the binocular lens. The lustre of pollucite is vitreous, sometimes more fatty than that of quartz. Cleavage absent. Hardness equals that of quartz. Specific gravity of pollucite of the Ungursay deposit, 2.89. On ignition evolves water and becomes opaque.

To distinguish pollucite macroscopically from the accompanying minerals is in no way simple, and methods for its field determination have not yet been worked out. Judging by its appearance, it may easily be mistaken for quartz (especially for quartz solid and transparent varieties), sugar-like albite, white beryl, amblygonite or fine-grained solid quartz-kukkeite aggregate usually formed upon petalite. Pollucite may be distinguished from other minerals by the following features: from quartz by a whiter colour; from ambly-

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gonite and albite, by the absence of cleavage and by semi-transparency; from beryl, by a fine-grained structure, and, finally, from quartz-kukkcite aggregate, by the absence of a silky lustre. The presence of a large number of grey and less frequently, of pinky and violet-coloured small veinlets is the most characteristic feature of pollucite, which permits it to be easily distinguished from nearly all other minerals. These veinlets are for the most part parallel to each other; on rarer occasions they intersect in all directions forming a network. Similar veinlets in pollucite have been reported by nearly all investigators ([a],[b]).

A microscopic examination carried out by the author permitted several types of the veinlets to be established:

1. Veinlets of very small spodumen grains of irregular shape containing the tiniest mirmekite inclusions of an unknown mineral. Analogous formations were described in 1939 by P. Quensel in pollucites of the Varuträsk deposit (Sweden). According to Quensel's data, the mineral which gives mirmekite inclusions in spodumen is a caesium analogue of spodumen, so-called caesium-spodumen.

2. Veinlets filled with various products of spodumen alteration (albite and eucryptite, albite and muscovite-cymatolite, halloysite).

3. Veinlets of fine-scaled lepidolite.

4. Veinlets of an aggregate of fine-scaled mica of the gilbertite type.

A chemical analysis of pollucite from vein No. 3 of the Ungursay deposit made by N. I. Vlodavez yielded the following figures:

SiO_2 —	48.08 per cent	. . . .	0.800
Al_2O_3 —	17.20 » »	. . . .	0.159
Fe_2O_3 —	0.32 » »	. . . .	0.002
CaO —	0.57 » »	. . . .	0.010
Li_2O —	0.42 » »	. . . .	0.014
Na_2O —	3.10 » »	. . . .	0.050
K_2O —	1.00 » »	. . . .	0.010
Cs_2O —	26.61 » »	. . . .	0.018
H_2O —	2.79 » »	. . . .	0.156

Total—100.04

$SiO_2 : R_2O_3 : R_2O : H_2O = 5 : 1 : 1.05 : 0.96 \simeq 5 : 1 : 1 : 1$

In pollucites of the Krasno-Kordon deposit the Cs_2O content is somewhat higher and attains 30—32 per cent according to K. A. Nenadkievich's data. Spectrum analyses of the Ungursay pollucite carried out by Professor S. A. Borovick and N. V. Lisunov showed, besides strong lines of Si, Al and Cs, medium-intensity lines of Na and weak lines of K, Li, Sr, Ca, Mg, Mn and Fe, and also traces of Be, Sn and Ga. Pollucite can easily be distinguished from all accompanying minerals by the total isotropy and a low relief, as shown by microscopic observations. Index of refraction of pollucite from the Ungursay deposit determined by the immersion method was found to be $N = 1.527 \pm 0.001$.

Upon microscopic observations pollucite has been found to be represented by strongly fissured cataclized excretions of an irregular shape, which are intensely etched by quartz and albite. In thin-sections of weathered specimens from the Ungursay deposit it may clearly be seen how an argillaceous mineral, possibly kaolinite, is developed upon pollucite. Typically enough, pollucite is encountered in veins containing rubellite and lepidolite only. In this particular case the latter may be distinguished by its peculiar weak violet coloration.

Quantitative spectrum analyses of these lepidolites carried out by Professor Borovick showed them to contain an increased quantity of Cs_2O, attaining several per cent. These fine-scaled lepidolites coloured light lilac and associated with rubellites and achroites may serve, therefore, as hints for pollucite prospecting.

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The finding of pollucite in industrial quantities in pegmatites of the Kalbin Range and the fact that wide circles of geologists have been informed of this finding enable us to emphasize the necessity of a revision of other analogous regions with respect to pollucite, in the first place the region of the Turkestan Range and of the Borchevochny Ridge.

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PETROGRAPHY

ON THE DISTRIBUTION OF GYPSUM IN ROCK TYPES OF THE PRODUCTIVE FORMATION OF THE AZERBAIJAN SSR

By L. V. PUSTOVALOV and A. D. SULTANOV

(Communicated by D. S. Beliankin, Member of the Academy, 8. I. 1946)

The fact that gypsum excretions are present in the productive formation of the Azerbaijan SSR is in no way novel and has frequently been mentioned in literature. No universally accepted interpretation of conditions of calcium sulphate accumulation in rocks of this age has, however, so far been advanced, though a solution of the problem would be of essential importance for a reconstruction of sedimentation conditions in the age of the productive formation.

Unfortunately, neither morphological nor crystallographic, mineralogical and lithological examinations of gypsum excretions occurring in rocks of the productive formation and its analogues have been conducted on a sufficient scale; the problem as to a preferential confinement of gypsum to various petrographic types of sedimentary formations also remains obscure.

It is required, therefore, to accumulate precise factual evidence capable of elucidating all questions, particularly those of genesis, related to the presence of gypsum among rocks of the productive formation

In this communication we discuss exclusively the distribution of gypsum scattered more or less regularly in various rock types of the stratigraphic complex which presents interest to us, leaving out altogether large isolated gypsum excretions («chip-like» gypsum, *etc.*) of distinctly epigenetic character which fill mostly fissures cutting deposits of the productive formation. 749 chemical analyses of hydrochloric extracts from rocks of the productive formation and its «analogues» have been used for the study of this problem; the analyses are derived from a paper on lithology of the productive formation prepared for publication by one of us (A. D. Sultanov). Quantities of gypsum present in various rocks have been determined by a corresponding recalculation of SO_3 content defined by chemical analysis. The presence of gypsum in rocks has also been ascertained by microscopic observations of thin-sections.

Gypsum has been found to be totally absent in rocks of four exposures (on the Akh-su River, at Geoglar, Mal. Horami, Baba-Sanan). In rocks of other exposures for which analytical data were available, average gypsum content is illustrated by the following figures.

For the total of 749 rocks examined the average SO_3 content equals 0.55 and that of gypsum, 1.18 per cent.

The distribution of gypsum is by no means uniform in rock types which have been distinguished in accordance with the classification scheme suggested in one of our previous communications ([1]).

In the figure is shown the classification scheme for areno-alevrite-argillaceous rocks. Dotted lines trace the fields of consolidated types. Circles, diameters of which correspond in the conventional scale to the gypsum con-

Exposures	Number of determinations	Average content of	
		SO₂ in %	gypsum in %
On the Alasan River	58	0.62	1.33
Ag-Dara	40	0.26	0.56
Djafar-abad	29	3.82	7.13
On the Gerdyman-chai River	60	1.04	2.23
Airan-Tekjan	171	0.53	1.14
Gilgin-chai	32	2.70	5.80
At the Masazyr lake	57	0.56	1.20

tent, and figures in the circles represent the average content of this mineral in each given type of clastic rocks. The diagram gives a clear notion of the gypsum distribution over rock types distinguished according to their granulometric composition.

The minimum average gypsum content (0.09 per cent) is observed in the case of alevrolites; on passing to more coarse-grained varieties (to the sand apex) the average gypsum content undergoes a regular increase and attains in sands 1.14 per cent. We are inclined to associate this fact presumably to a
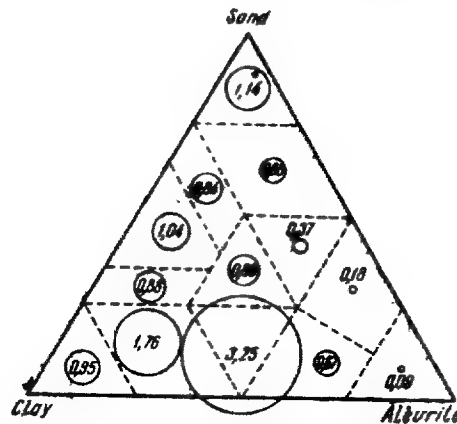
travel and increased evaporation of mineralized solutions in capillary pores at outcrops of more coarse-grained arenaceous rocks which,for obvious reasons, possess relatively larger pores, this favouring intense evaporation of a mineralized moisture, a supply of fresh portions of water to the surface of outcrops from below and, consequently, an epigenetic accumulation of gypsum, for the most part in those particular rocks (all specimens have been taken from natural exposures or from shallow clearings).

The distribution of gypsum in alevrite-argillaceous rocks is of considerable interest. On passing from alevrites to clays an increase in the content of argillaceous admixtures is accompanied at the beginning by an increase in the average gypsum content, and the latter attains its maximum in loams (3.25 per cent), but a further increase of grain fineness brings about a drop in gypsum content which in clays proper equals but 0.95 per cent. If gypsum excretions were of syngenetic character, we would be justified in expecting them to be preferentially associated with the most fine-grained sediments, since the latter accumulate under conditions of the most retarded or even totally lacking inflow of waters, i. e. under the conditions which favour a precipitation of gypsum.

Therefore, a preferential association of gypsum not with pure argillaceous rocks but with loams may be considered as an indirect indication of the

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epigenetic character of this mineral in argillaceous rocks. This conception is in good agreement with our field observations carried out in various localities of the Azerbaijan SSR in summer 1945, when epigenetic gypsum excretions were found in abundance, whereas no gypsum excretions were observed which might even supposedly have been considered as syngenetic excretions. Supplementary observations are required, however, if we attempt to solve the problem which has now arisen, *viz.* to establish the factors responsible for the maximum confinement of gypsum to loams rather than to any other type of clastolytes.

In various types of poorly sorted out chlidolite rocks the average gypsum content was found to be approximately uniform and to range from 0.84 to 1.04 per cent.

Though a considerable wealth of analytical data has been involved in the determination of gypsum distribution among rock types, it supplies us at present with no direct indications as to the syngenetic character of this mineral in sediments of the age of the productive formation.

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GENETICS

EXPERIMENTAL PRODUCTION OF INTERSPECIFIC HYBRIDS IN THE GENUS *QUERCUS*

By S. S. PJATNITZKY

(Communicated by A. A. Richter, Member of the Academy, 15. XII. 1945)

The first attempt of obtaining experimentally interspecific hybrids in the genus *Quercus* was made in 1845 by Klotsch ([1]) who crossed *Q. Robur* Wild. (*Q. sessilis* Ehrh.) with *Q. pedunculata* Ehrh. and obtained a hybrid which, at the age of eight years, proved to be by 1/3 taller and more vigorous than the plants of the same age belonging to the parental species. Gale ([2]) attempted unsuccessful experiments on crossing *Q. falcata* Mchx. with *Q. Phellos* L., with the purpose of studying the question as to the origin on *Q. heterophylla* Mchx. which was supposed to be of a hybrid nature. Geschwind([3]) obtained 4 hybrid plants from the cross *Q. sessilis* Ehrh. with *Q. pedunculata* Ehrh. Ness ([4],[5],[6]) reported on the production of hybrids *Q. virginiana* Mill. × *Q. lyrata* Walt., *Q. virginiana* Mill. × *Q. bicolor* Willd., *Q. virginiana* Mill. × *Q. minor* Sarg., *Q. virginiana* Mill. × *Q. platanoides* Ludw. The total number of hybrid plants produced by the author named above attained 20. The F_2 grown from these plants was described by Ness himself and by Jarnell ([13]). Kolesnikov ([7]) was engaged in work on interspecific crosses in oaks, and described the production of a rapidly growing hybrid between *Q. Robur* L. and *Q. macrocarpa* Mich.

Our work on interspecific hybridization in *Quercus* was started in 1937 with the aim of producing hybrids between species far remote both taxonomically and geographically. The technique of the work was described by us in an earlier note ([8]). The work was interrupted by the war. The majority of the hybrid progeny survived and is developing now. Two hybrid forms have been described by the present author ([9]) by the names: × *Q. Miczurinii* and × *Q. Wyssotzkyi*.

In the table are presented data on five-year long experiments (1937—1941).

During the five years of work a total of 39 hybrid combinations were produced; negative results were obtained in 8 instances only. We succeeded in producing hybrids between species belonging to distinct sections of the genus; the respective crosses yielded in certain instances a rather high proportion of hybrids. The following crosses proved to be a high success: *Quercus macranthera* F. et M. (sect. *Mesobalanus*) with *Q. alba* L. and with *Q. macrocarpa* Mich. (sect. *Lepidobalanus*). It has been found that: 1) no hybrids are ever produced from pollinations of *Q. borealis* v. *maxima* (sect. *Erythrobalanus*) with the pollen of species belonging to other sections; 2) *Q. macrocarpa* and *Q. montana* when pollinated with pollen of other species produce in the majority of cases but a few hybrid offspring; 3) in individual crosses the success varies as to the direction of the cross.

The incrossability of *Quercus borealis* v. *maxima* with other species when this species is used as the female parent is explained by its belonging to sect.

♀ plant	♂ plant	Number of flowers pollinated	Number of hybrid acorns obtained	Number of plants grown which survived by Oct., 1945	Best results of crossing in individual exper. % to number of flowers pollinated
Q. Robur L.	Q. macranthera F. et M. . .	4 119	41	25	2.0
	Q. macrocarpa Mich.	3 130	24	6	2.7
	Q. borealis Mich. v. maxima Asch. . . .	5 088	21	5	3.2
	Q. montana Willd.	1 024	42	1	8.4
	Q. alba L.	4 709	24	4	0.9
	Q. Suber L.	8 719	38	3	1.4
	Q. fastigiata Lam.	322	84	12	13.0
	Q. sessilis Ehrh.	749	0	—	0
	Q. Robur L.	2 689	117	24	6.0
Q. macranthera F. et M.	Q. Robur L.	7 916	259	134	10.4
	Q. macrocarpa Mich.	3 990	780	91	45.0
	Q. borealis Mich. v. maxima Asch. . . .	4 643	56	12	6.5
	Q. montana Willd.	4 648	15	1	0.6
	Q. alba L.	5 194	1 019	68	46.0
	Q. Suber L.	6 442	15	7	0.7
	Q. macranthera F. et M. . .	3 009	64	37	0.15
Q. macrocarpa Mich.	Q. Robur L.	3 955	1	1	0.13
	Q. macranthera F. et M. . .	8 732	32	17	2.0
	Q. borealis Mich. v. maxima Asch. . . .	2 466	6	2	0.5
	Q. Suber L.	3 411	0	—	0
Q. borealis Mich. v. maxima Asch.	Q. Robur L.	4 787	1	—	0.15
	Q. macranthera F. et M. . .	1 314	1	—	0.2
	Q. macrocarpa Mich.	1 817	0	—	0
	Q. montana Willd.	858	2	—	0.2
	Q. alba L.	1 598	0	—	0
	Q. Suber L.	3 094	0	—	0
	Q. borealis Mich. v. maxima Asch. . . .	81	0	—	0
Q. fastigiata Lam.	Q. Robur L.	843	39	4	4.0
	Q. borealis Mich. v. maxima Asch. . . .	428	1	—	0.3
	Q. macranthera F. et M. . .	369	1	—	0.3
Q. montana Willd.	Q. Robur L.	1 722	8	—	0.5
	Q. borealis Mich. v. maxima Asch. . . .	1 615	0	—	0
	Q. Suber L.	240	0	—	0
Q. Suber L.	Q. Robur L.	1 482	7	—	4.2
	Q. macranthera F. et M. . .	371	1	—	1.7
	Q. macrocarpa Mich.	1 887	35	—	21.2
	Q. montana Willd.	269	3	—	5.4
	Q. alba L.	1 170	6	—	8.3
	Q. borealis Mich. v. maxima Asch. . . .	1 374	8	—	2.5
	Total	105 274	2 695	454	—

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Erythrobalanus which is characterized by a biennial period of maturation of acorns. According to Hofmeister ([10]) and Conrad ([11]), fertilization takes place in such species within 13—14 months after pollination, while in those with annual maturation of acorns this interval is 2—3 months only. The incompatibility of the life cycles of the pollen of the species showing annual maturation of acorns with the course of development of the ovary in species with a biennial maturation may be the cause of the incrossability of these species. The poor results of crosses between *Q. macrocarpa* and *Q. montana*, even in the cases when pollinated with pollen of systematically related species, should be put down to the effect of the individual properties of the trees pollinated. The limited number of blooming trees of these species available made it impossible to select individuals most suitable for the pollination with pollen of other species than those to which they belonged. Experience showed that the phenomenon referred to above actually occurred in nature, for trees of the same species growing in close vicinity yielded different results when cross-pollinated.

The different results observed in direct and reciprocal crosses make one suppose that in interspecific crosses in oaks the results depend not only on the biological compatibility or incompatibility of the gametes, but also on such agencies as develop in the course of the long period between pollination and fertilization.

The best results of crossing were observed when the pollen was applied to the stigmas 3—4 days after the beginning of blooming, and in case only such pollen was used as gave no less than 50 per cent germination *in vitro*.

The interspecific hybrids obtained showed in the majority of cases mixed characters and were morphologically somewhat heterogeneous in the first generation. The hybrids obtained by crossing *Quercus Robur* L. exhibited a great influence of this species, especially in combinations where it served as the female parent, which could be explained after Michurin ([12]) by the effect of the developmental conditions to which the common oak is better adapted than other species are.

The hybrids of *Quercus* are of some economical interest owing to the heterosis exhibited by them, as well as to the fact that some of the hybrids of cork-oak proved to be fully winter-hardy, which made them capable to survive under the climatic conditions of the northern part of the Ukrainian steppe, where they are now cultivated.

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GENETICS

SOMATIC REDUCTION AS A CAUSE OF A PECULIAR MOSAICISM IN TETRAPLOID BUCKWHEAT

By V. V. SAKHAROV

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The autotetraploid form of buckwheat differs both morphologically and physiologically from the ordinary commercial buckwheat from which it has been produced experimentally by simply doubling the chromosome number. The new form does not cross, moreover, with the original either under field conditions or in especially arranged experiments.

The above circumstances, along with karyological distinctions form a basis for raising the new form of buckwheat to a specific rank. The history of its independent existence, which dates of quite a recent epoch and which, we hope, will succesfully continue, permits the new form to claim the rank of a species.

Physiological isolation which is granted to us by Nature at the very moment of the creation of an autotetraploid form is not merely of theoretical interest. It bears, moreover, one of the most important factors of the stabilization of the new form in the fields. The uncrossability of the tetraploid form with the ordinary one precludes the appearance of unbalanced triploids and, therefore, a mass regression of the young population to the initial diploid condition. This stability of the new form which is fertile only *inter se* enables us to foresee that it will prove constant under the hands of agricultural practitioners.

A detailed study of the new form went to discover a rather unexpected phenomenon, however. When the plants were thrashed individually and the seeds that ripened on each carefully counted, there were found, although very scarce, exceptional small seeds. For the first time such observations were made in the summer of 1942. Any possible error was precluded by the fact that on the second year of our work only such plants were preserved as were cytologically checked and all the diploid individuals were discarded. In the following years we continued recording such deviations, and their list became considerably wider.

The exceptional small seeds repeat the coloration and the pattern of their gigantic sibs (these characters vary greatly from plant to plant), differing from them in size only. Below are cited only three typical instances of aberrating plants found in the fields of 1944.

Plant No. 19-60: among 476 large seeds (average weight of one seed 34.0 mg) 8 small seeds (average weight of one seed 23. 7 mg)

Plant No. 21-75: among 548 large seeds (average weight of one seed 43.2 mg) 4 small seeds (average weight of one seed 22. 5 mg)

Plant No. 22-30: among 794 large seeds (average weight of one seed 35.9 mg) 5 small seeds (average weight of one seed 18.0 mg).

These differences in the average weights of the large and small seeds correspond exactly to those between tetraploid and diploid seeds.

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It remained to find out whether the small seeds were not accidental purely phenotypic aberrations caused by some shortage in the plant's development. In order to do this the «exceptional» seeds were germinated, and a number of cytological analyses for the chromosome number were made. Indeed, some of the small seeds developed tetraploid 32-chromosome seedlings (*i. e.* proved to be phenotypic deviations), but in a number of cases the seedlings undoubtedly were diploid forms possessing the standard chromosome number equal to 16.

The occurrence of diploid seeds on tetraploid plants demanded investigation of the distribution of the phenomenon in question. It appeared that among a material belonging to a number of families 113 plants out of 2216 (5.1 per cent) gave small «exceptional», seeds among a vast majority of large ones. Out of the total of over half a million (536 143) seeds collected from all the plants studied there were only 369 small seeds, which makes 0.069 per cent only.

It should be noted that this moderate proportion of «exceptional» small seeds is in reality greatly exaggerated, as the majority of small seeds are, as a matter of fact, nothing else than underdeveloped tetraploid ones.

In analysing the causes of the appearance of diploid seeds upon tetraploid plants, we could at first only hazard suppositions. The first of them was to admit the possibility of parthenogenetic development of seeds from unfertilized egg-cells or from other cells with a reduced chromosome number. The other hypothesis which seemed to us less probable was to suggest that there was a possibility of the appearance of somatic regions with reduced chromosome number and of a subsequent formation of generative organs out of them. Subsequent new observations carried out in 1945 made us decide in favour of the latter hypothesis.

When the tetraploid plants were carefully examined, on some of them, which were very few in number, were found peculiar mosaic spots. This mosaicism usually occurred on fully healthy, normally blooming and fruiting plants. Usually there could be seen a sharply limited spot upon one of the leaves, occasionally close to the middle, in other cases closer to the edge, or even at the very edge of the leaf blade. These spots were distinguished from the surrounding tissue by their even and light coloration, and by a kind of impression. They remained unaltered without increasing or changing their outlines. The areas covered by these spots did not exceed 3—5 per cent to the whole area of the leaf blade and were often considerably smaller, looking as scarcely perceptible light stripes.

We dispose of no precise data on the distribution of this mosaicism, but at our experimental plots (each containing 500 plants) we have found up to ten and even more plants possessing at least one spot. It is of interest that some of the plants bore several spots, all of them being sometimes gathered on a single leaf.

A most careful search for such a mosaicism in diploid plants proved unsuccessful.

The discovery that this mosaicism was observed in tetraploids only made us associate it with the observations on the distribution of the «exceptional» small seeds of the tetraploids. Our principal working hypothesis accounting for the two phenomena became the assumption that isolated patches of diploid tissues might arise in the bulk of the tetraploid tissue.

In favour of the above hypothesis on the nature of this mosaicism was the colour of the spots, which was identical with that of the leaves in diploids and considerably paler than that in tetraploids. A smaller thickness of the mosaic spots as compared with the surrounding tissue seemed also to speak in favour of this idea.

In collaboration with S. I. Bekshi, undergraduate of the Voronezh University, we started a comparative cytological investigation of the light and dark areas of the upper epidermis of the same leaf blades. With the aid of a

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camera lucida the outlines of the light and dark areas belonging to the same leaf were drawn on paper. The maximum diameters of the cells were measured on the drawings.

Data on 10 mosaic leaves have shown in 7 instances statistically significant differences in the sizes of the cells in favour of the darker areas. Only in 3 cases it was found impossible to prove the significance of the difference, although the cells of the darker areas were greater, too, than those of the light ones.

Summarized data of the measurements of the cells of the dark and light areas for all the ten cases are as follows:

$$M \text{ of the } \text{ dark area} = 1.670 \pm 0.016 \ (n = 686)$$
$$M \text{ of the light area} = 1.408 \pm 0.013 \ (n = 870)$$

$$M_{\text{diff}} \qquad\qquad = 0.262 \pm 0.022$$

The difference found from the summarized material is 11.7 times its standard error, *i. e.* absolutely significant.

When the cell size of the light areas of tetraploid leaves was compared with that of the leaves of normal diploid buckwheat, no real difference was found in accordance with expectation.

The crucial evidence in favour of the theory that the mosaic areas were in fact diploid patches in the tetraploid tissues might be furnished only by direct chromosome counts; however, there could be found no dividing cells in the epidermis of the leaves. Even when this difficulty could be overcome by studying growing cones, one could not foresee where and how to look for the initial spots, where mosaic areas would subsequently arise.

When discussing the problems connected with the present study with S. L. Frolova, she informed me that, in the light of the hypothesis here advanced, some karyological observations which seemed obscure before, now acquired a definite meaning. While studying on aceto-carmine slides root-tips of tetraploid plants of buckwheat, she repeatedly observed among the multitude of 32-chromosome cells individual cells possessing diploid chromosome complement consisting of 16 elements.

Having studied the occurrence in tetraploid plants of exceptional diploid seeds and peculiar mosaic spots and compared the two phenomena with each other, we arrived at a conclusion that a common cause must be responsible for the two phenomena. It lies in an unusual mode of division of individual tetraploid cells, leading to the formation of diploid daughter cells, for lack of a previous chromosome splitting. We believe that we are facing here an entirely specific phenomenon of «somatic reduction», which occurs in a historically youthful form which had no time as yet to adapt itself to its new tetraploid condition.

In introducing the term «somatic reduction» we are fully aware of the responsibility. The conclusion cannot be finally adopted until the very mechanism of somatic reduction is borne out in a way similar to that shown for cell division in the metamorphosis of mosquito ([1]) and in dividing cells of the polyploid tissues of the mutant *cancroidea* in *Antirrhinum* ([2]). In our special studies of the somatic mitoses in tetraploid buckwheat we expect to find instances of «somatic conjugation» of the chromosomes, although they may be rare. Otherwise one cannot imagine regular disjunction of the pairs of the fourfold set of the eight chromosomes of the tetraploid cell.

In concluding a few words are to be said as to whether the appearance of accidental diploid seeds may affect the practical value of the new form. The admixture of individual diploids to great massives of tetraploids cannot lead to any shift of the whole population towards diploids. To account for such a shift one should assume a certain advantage of diploids in the struggle for life, which has never been noticed in any respect. Moreover the fact that single diploid plants would be scattered among the multitude of tetra-

ploid ones would be highly unfavourable with regard to their pollination, since the pollen of tetraploid is incompatible with them. Besides, the tetra-ploids mature somewhat later than diploids (5—10 days) and, which is more important, do not shed seed when mature, so that they can remain in the field until fully mature; this would again place diploids in an extremely unfavourable position, as they would shed seeds while standing in the field. Finally, even in the case of an undesirable propagation of diploids taking place, it would be far simpler to control them than many weeds. It will suffice to fit a sieve, and the small seeds of diploids will be eliminated readily and at once.

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EFFECT OF ACENAPHTHENE UPON NON-DISJUNCTION OF SEX CHROMOSOMES IN *DROSOPHILA MELANOGASTER*

By M. O. ZIVIN

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The effect of the alkaloid colchicine upon the nuclear reaction in plants was studied both abroad and in this country [1-4]. This alkaloid has been found to cause in many plants irregularities of mitosis, leading to polyploidy. In their attempts to use colchicine for the treatment of cancer, Amoroso [5] and others injected colchicine solutions to animals bearing tumours; it was found that the named alkaloid caused in some of the animals regression and even disappearance of the tumours. Other workers [1-3] found that colchicine produced very strong effect upon karyokinesis, which made it possible to regard it as a strong karyoclastic poison.

The toxicity of colchicine, as well as the scarcity of its supply encouraged us to the search for other chemical substances up to replace it. While studying the effect of acenaphthene and colchicine Shmuck [6] found that the former acted in a manner similar to that of colchicine, though somewhat stronger. The characteristic feature of the effect of acenaphthene consists in the disturbance of the normal process of cell division, leading to polyploidy. Besides, acenaphthene is free of the toxic property of colchicine, and its effect upon the plant cells is considerably milder. It might be anticipated that acenaphthene would have a similar effect upon animal cells, including cells of tumours.

Navashin [7] studied the effect of acenaphthene upon the division of the cell nucleus in various plants and found that acenaphthene, similar to colchicine, caused variation in chromosomal composition, the main phenomena observed being somatic haploidy and somatic segregation. Dontcho Kostoff [8] in his studies arrived at a conclusion that histological and cytological changes in tumourous tissues of plants were identical with those occurring in cancerous tissues of animals. Shmuck, Gusseva and Iljin [9] studying alterations in plant tissues arrived at a conclusion that the effect of acenaphthene upon the plant cells may be compared to that of cancerogenic substances upon animal cells.

Data obtained in the laboratory of Shabad on the effect of acenaphthene upon graftable carcinoma and sarcoma in mice point to a tendency to inhibit the growth of these tumours [10]. All the students arrived at a conclusion that acenaphthene is much similar to colchicine as to its effect upon the cell nucleus and, at the same time, is free of the toxic effect of the latter substance. Inasmuch as the molecular structure of acenaphthene represents in a sense a part of the molecule of cholantrene, the latter being notably a cancerogenic substance, the question arises as to the nature of the effect of acenaphthene upon animal cells.

H₂C — CH₂
Acenaphthene

H₂C — CH₂
Cholantrene
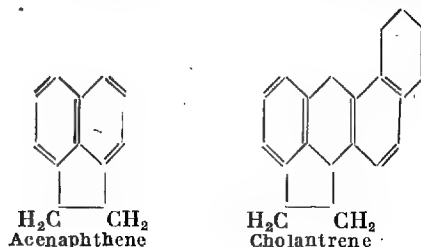
The aim of the present study was to find out the influence produced by acenaphthene upon the chromosomal apparatus of *Drosophila*. The primary non-disjunction of chromosomes may readily be discovered in the cross ♀ white × ♂ Normal. Flies of the white line were fed on food with acenaphthene added at varying concentrations, in the form of alcoholic solution or of powder. The amount of food in each vial was always equal to 5 cm³. After the deposition of eggs the flies were removed. White females which had developed on such a medium were transferred to usual food and crossed with males of the Florida line. The table shows the numbers of the exceptional flies in the experiment and in the control.

Exper. No.	Number of flies examined	♂ normal	♀ white	inter-sexes	others	total number	%	Remarks
65	2 710	—	1	1	2 mosaics	4	0.15	Mosaics: 1 ♀ left eye white; right eye red; 1 ♂ left eye red. right eye white
66	1 479	3	1	6	—	11	0.74	
67	3 184	3	—	1 8 1	—	12	0.38	
68	1 832	3	1	2	—	6	0.33	
89	733	—	1	2	—	3	0.41	
90	851	2	—	3	—	5	0.59	
91	325	—	—	1	1	2	0.61	
92	561	1	—	2	—	3	0.53	
Total . .	11 675	12	4	27	3	46	0.4 ± 0.06	
Control .	8 292	—	—	—	—	—	0.0	

In order to find out whether there was primary non-disjunction, exceptional females F_1 were crossed with males of the Florida line. In the progeny many exceptional individuals were obtained. Red-eyed males proved sterile. From a study of the effect of acenaphthene upon *Drosophila melanogaster* the conclusion should be drawn that acenaphthene causes non-disjunction of the sex chromosomes.

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PLANT PHYSIOLOGY

ROOTING OF CUTTINGS OF *QUERCUS SUBER* L. TREATED WITH CHEMICALS

By D. A. KOMISSAROV

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Vegetative propagation of cork-oak by means of cuttings may be turned to advantage in the first place when breeding for new highly productive varieties.

It has been shown by several authors, however (Hubert and Beke ([2]) and others), that the cuttings of cork-oat fail to root, either when planted in the usual way, or if pre-treatment with synthetic growth substances is applied. The author's first experiments, carried out at Sochi and Sukhumi in 1937—1938 gave no favourable results either: cuttings taken from plants 2—3-year-old and treated with α-naphthalene acetic acid could be made to root in no more than 5 per cent of cases, while those treated with heteroauxin failed to root at all. In the summer of 1941 the author carried out a series of experiments in Leningrad, under quite different conditions, and obtained this time a rather high per cent of successful rooting.

M a t e r i a l a n d M e t h o d A p p l i e d . Acorns were planted into boxes in 1939. During the first four months upon planting the seedlings were grown in a room, under continuous electric illumination. During spring and summer of 1940 the plants were kept in the open; at the approach of autumn they were transferred to a greenhouse, where they passed winter at a temperature of 3—5°C. By the time of cutting, in summer 1941, the two-year-old plants have already attained a height of 80—90 cm.

The cuttings were made 5—6 cm long, bearing 3—5 intact leaves with well-developed, uncorked lateral shoots of the first order of branching.

As chemicals that would stimulate the cuttings to form roots were tested: heteroauxin, α-naphthalene acetic acid, and also such substances as are no growth stimulants, *viz.* anthracene, anthranilic acid and nicotine. According to data reported by Traub ([3]), the three last substances had been found to produce upon the rooting of cuttings of *Passiflora quadrangularis* and *Bignonia venusta* a stimulating effect similar to that of β-indolbutyric acid. In addition to the chemicals listed above, we also tested two new substances, *viz.* nicotine-sulphate and anabasine-sulphate.

All these substances were applied to the cuttings either in the form of dry powders mixed with talc ([1]), or as water solutions ([4]).

The powders were mixed at a ratio of 3 parts of the respective active substance to 1000 parts of talc. The mixture was prepared in the following way. 15 mg of the substance to be tested was dissolved in 2 cm³ of 96° alcohol and then thoroughly mixed in a porcelain mortar with 5 g of talc. The mixture was dessicated at 16°C during 20 hours, and then reduced to a fine powder.

The bottom ends of the cuttings were immersed for a moment into the powder to a depth of 1 cm. The cuttings were carefully freed from the excess of powder that stuck to them by slightly tapping at the brim of the vial. For

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sake of comparison, control cuttings were treated with pure talc powder. The cuttings treated in this way were planted directly into glazed culture-boxes. It should be pointed out that the tips of the cuttings were slightly moistened before immersing them into the powders, for powders fail to adhere to dry cuttings, whereas excessively moist ones would have too much powder stuck to them, which might do them harm. For this purpose the cuttings were usually kept for some time with their lower tips immersed in water. Previous to planting the excess of water was eliminated by slightly shaking the cuttings, and by gently applying the fresh cut to a sheet of blotting paper. The concentrations of the aqueous solutions of the substances tested and the time of treating the cuttings with each of them are given in Tables 1, 2 and 3.

Table 1

Experiment Carried out on June 26, 1941

Cuttings treated with dry pulverized substances	Number of cuttings	Rooting of cuttings, in %	Shoot formation, in %	Number of cuttings with calluses, in %	Time of rooting, in weeks	Mean length of roots per l cutting, in cm
Control (pure talc). . . .	20	20	0	80	8	8
α-naphthalene acetic acid, 3 parts to 1000 parts of talc	20	90	90	100	6	32
Heteroauxin, 3 parts to 1000 parts of talc	20	70	20	100	7	30
α-naphthalene acetic acid, 1 part+heteroauxin, 1 part+anthracene, 1 part to 1000 parts of talc . .	20	100	50	100	6	33
Anthranilic acid, 3 parts to 1000 parts of talc . .	20	40	0	100	8	18
Anthracene, 3 parts to 1000 parts of talc	20	20	0	90	8	13

Table 2

Experiment Carried out on June 26, 1941

Cuttings treated with water solutions	Number of cuttings	Rooting of cuttings, in %	Cuttings which developed shoots, in %	Number of cuttings with calluses, in %	Time of rooting, in weeks
Control, water during 12 hrs . . .	20	40	0	90	8
Heteroauxin 0.01%, 12 hrs	20	50	10	100	6
ɒ-naphthalene acetic acid 0.01%, 12 hrs	20	70	30	100	6
Anthracene, saturated solution in 0.5% alcohol, 12 hrs	20	50	10	100	7
Anthranilic acid 0.05%, 12 hrs . .	20	60	10	100	7
Anthranilic acid 0.10%, 12 hrs . .	20	60	0	100	7
Mixture consisting of: heteroauxin 0.02%, α-naphthalene acetic acid 0.02%, anthracene, satur. solution, taken at equal parts, 12 hrs . .	20	80	30	100	6
Same mixture, minus anthracene, 12 hrs	20	80	30	100	6
Anabasine-sulphate, 0.02% (concentration to base) 12 hrs	20	60	0	100	7

Table 3

Experiment Carried out on July 30, 1941

Cuttings treated with water solution	Number of cuttings	Rooting of cuttings, in %	Shoot formation, in %	Number of cuttings with calluses, in %	Time of rooting, in weeks
Control, water during 6 hrs	20	20	0	100	9
Nicotine 0.02%, 6 hrs.	20	40	0	100	8
Nicotine-sulphate 0.02% (concentration to base), 6 hrs.	20	70	0	100	8
Anabasine-sulphate 0.02% (concentration to base), 6 hrs.	20	60	0	100	8
Heteroauxin 0.01%, 6 hrs.	20	40	0	100	8
a-naphthalene acetic acid 0.01%, 6 hrs.	20	80	0	100	7

The boxes which served for the planting of cuttings were $50 \times 50 \times 28$ cm in size. In the bottom of each box holes were pierced, 1.5 cm in diameter, at 6 cm intervals. The floor of the box was covered with a layer of brick-rubble 6 cm thick, for drainage; this was overlain by a 6 cm layer of light sandy loam; and finally by a 3 cm layer of washed quartz sand.

The box was covered from above by a pane of clean glass. The boxes were put on shelvings in a greenhouse. In order to secure better aeration, free space was left between the bottom of the box and the shelving. The panes of the greenhouse were kept clean throughout the experiments.

To preserve the cuttings from direct sun rays, the boxes were screened by means of a shingle-shield fixed above them at a height of 1.5 m, with small windows made in it, which were covered with double gauze.

Owing to this arrangement the cuttings received moderate scattered light. The midday temperature of air within the boxes during June, July and the first 10 days of August was maintained at about 32—34°C, with a variation range from 27° at 8 a. m. to 30° at 8 p. m. Watering was done with the aid of a pulverizer twice a day; each box received about 100 cm³ water at once. Thus watered, the cuttings suffered neither from excessive, nor from deficient humidity. The conditions established in the culture-boxes were favourable for the rooting of cuttings.

In agricultural practice similar conditions may readily be established in above-ground hothouses supported by low pillars, arranged in accordance with the description of culture-boxes given above.

R e s u l t s . From the results of our experiments, as summed up in Tables 1—3, it may be seen that under favourable conditions control cuttings taken from 2-year-old cork-oat plants were found to root in 20—40 per cent of the cases, while those subjected to pre-treatment with chemicals rooted in 70—100 per cent of cases.

From among the chemical substances applied as rooting stimulants, the most highly efficient proved the a-naphthalene acetic acid; next followed heteroauxin, anthranilic acid, and also the new substances tested by the author for the first time, *viz.* nicotine-sulphate and anabasine-sulphate.

Because of their cheapness and availability the two last substances deserve to be put to test on other plant-species, as well. Anthracene was found to produce but a very slight influence upon the increase in rooting-rate of the cuttings treated with it.

When applied to the cuttings in the form of dry powders mixed with talc, a-naphthalene acetic acid and heteroauxin were found to yield somewhat better results than the water solutions of those substances. Especially efficient proved a mixture consisting of 1 part of naphthalene acetic acid, 1 part

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of heteroauxin and 1 part of anthracene per 1000 parts of talc: In this case 100 per cent of successful rooting were obtained.

From among the substances tested for their stimulating effect upon the development of the root system, especially efficient were found to be α-naphthalene acetic acid and heteroauxin (Table 1). In cuttings treated with these substances, roots were set somewhat above the callus, laterally, from under the cortex, through the formation anew of root rudiments. In individual cuttings the roots appeared from the callus, which occurred some 4—5 weeks later than from under the cortex.

When propagation by cuttings was undertaken at an early date (26 June) all the cuttings which had been treated locally with dry pulverized α-naphthalene acetic acid mixed with talc applied to the cut, were found to develop by the end of summer shoots up to 15 cm long, whereas the control cuttings had all their buds still quiescent (Table 1).

In another experiment (Table 2) carried out on the same day, cuttings treated with a 0.01 per cent aqueous solution of α-naphthalene acetic acid were found to set root in 70 per cent of the cases, but only 30 per cent of these developed shoots, which occurred by three weeks later than in cuttings treated with dry powders.

When cuttings were prepared at a later date (30 July), the cuttings treated with water solutions of α-naphthalene acetic acid, heteroauxin and other substances were likewise found to root well, but their buds remained in a state of rest, similar to those of the controls (Table 3).

We thus see that when given at definite optimal doses, α-naphthalene acetic acid may under favourable external conditions stimulate not only rooting of the cuttings treated, but also the development of buds and shoots in these cuttings.

The results of the present investigation may prove of use in the breeding of cork-oack, as a means to accelerate propagation by cuttings of 2—3-year-old plants, grown from acorns produced by the most valuable plants.

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PLANT PHYSIOLOGY

MODE OF GROWTH OF LEAVES IN SUMMER-WHEAT AS INFLUENCED BY NITROGEN DIET

By B. A. CHIZHOV

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Investigations carried on at the Institute for the Study of Grain Crops of the South-East went to demonstrate the expedience of applying to wheat grown under irrigation adjustable fertilization systems. Each system is supposed to consist of two links. One of them is constant, the fertilizers (which are largely phosphorus and potassium preparation) are introduced previous to planting; the other link is adjustable, the fertilizers (mostly nitrogenic ones) being applied in the course of the plant's development, in accordance with its demand which may vary as to the conditions of growth. When choosing the composition of the nutrient substances and settling the dates of their administration in the second link of the system, it was found necessary to work out such methods as would permit in a most simple and rapid way to evaluate the need of a given crop for nutrient substances according to the stage of development. Among the variety of procedures commonly applied in agricultural practice, the most promising seemed to be those based upon external characters ([5-9],[11],[12]).

With the aim to shed more light upon this problem, the author undertook in 1944 experiments on the influence of nitrogen food on the length and breadth of leaves, number of adventitious stalks and colour of leaves. Of special interest are the results of the investigation of the growth of leaves as affected by nitrogen diet. The most essential of them are discussed in the present note.

The growth of leaves in summer-wheat was studied by number of authors ([1-4],[10]). Particular interest, especially in arid regions, belongs to investigations dealing with the influence upon the development of leaves of the varying supply of the plants with water ([1],[4]). The influence of variations in the mineral food upon the growth of leaves has so far not been studied in sufficient detail. To the author's knowledge, the information available do not go beyond a paper recently published by Gushchin ([4]).

As material for our experiments served soft summer-wheat, variety *Lutescens* 758. The plants were grown in sand cultures with Helriegel solution, in soil-sand cultures and on micro-plots, in the field. The various experimental sets served to create a different supply of the plants with nitrogen. In sand cultures the leaf blades of the different storeys were measured each second day; in soil-sand cultures, twice a week; in the field experiment, once a week. Along with the study of the growth of leaves, we also made attempt to determine the length of their life-activity. The life-activity of a given leaf was conventionally termed as «stopped» whenever a large part of the leaf blade was found to have lost its green colour, and began to dry up. On account of the data obtained, the growth and life-activity of leaves in wheat and the influence produced upon these processes by nitrogen food, are characterized by the following peculiar features.

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According to the ideas developed in the works by Krenke and his pupils, a peculiar cyclicity is observable in the growth and life-activity of plants. The life of plants is accompanied by continuous processes of ageing and rejuvenescence. This cyclicity in the plants' development, which has been borne out by Krenke, is clearly manifest both in the growth and life-activity of leaves of wheat which are regularly appearing and dying off. The open development of summer-wheat is characterized by a simultaneous pairwise growth, and sometimes, by consecutive growth of single leaves (in the upper storeys). Two leaves make their growth at the same time: 1 and 2, 2 and 3, 3 and 4, and so on. As the lower leaves are ageing and stop to grow, they are substituted for by young leaves,situated higher upon the stalk. The period of growth of leaves at the different storeys varies in soft wheat (*Lutescens* .758) within a range from 7 to 12 days, and rarely beyond these limits. The time interval between the appearance of two adjacent leaves is somewhat shorter, and this is why the leaves are growing largely pairwise.

The periods of life-activity and work of the leaves is much longer than the period of growth. For the leaves of the bottom storey these periods are equal to about 20—30 days, while for the uppermost internode they attain 50—60 days.

The number of leaves at work exceeds that of growing leaves. According to the author's observations, simultaneously at work are leaves 1 and 2, and partly, 3; 4 and 5, and partly, 3 and 6; 7 and 8, and partly, 6. Under the influence of a liberal nitrogen diet at the beginning of vegetation and especially in the case of additional rations of nitrogen fertilizers, the periods of life-activity and work of the upper leaves are protracted owing to the rejuvenescence of the plant, and may attain 60—70 days.

The periodicity observed in the development and life-activity of the leaves of different storeys is observed in each individual leaf, as well. The period of life of a leaf can be subdivided into three stages, each of which is characterized by peculiar morphological and biological features of its own. The first stage, that of growth, when the leaf is working largely to provide for itself. The second stage, that of the most intense activity, when the leaf is working both to provide for itself and for other organs. The third period is that of gradual ageing and of a rapid drop in the life-activity, when the leaf ceases to work for itself, and provides for other organs only. If we compare the various stages of the life-activity of leaves with the stages of growth and periods of formation of the individual organs, we get an idea as to the special trend in the work of leaves belonging to the different storeys. If made to fit within a rough scheme, the work of the three bottom leaf storeys is spent on the formation of both vegetative and generative organs, *i. e.* of new leaves, adventitious stalks, hairs and florets of the ear rudiment. The work of the leaves of the middle storey is spent on making those organs grow, while the work of the leaves of the upper internode secures the filling of the kernels.

The length, breadth and shape of leaves, as well as the trend of their work, are subject to variation under the influence of various external agencies. The leaves of the individual storeys of summer-wheat are highly susceptible to the varying conditions of nitrogen supply. In the case of liberal nitrogen supply, the rest of conditions being favourable for the growth of the plant, the plants usually show an increase in the length of their leaves from the bottom storey to the subsequent ones (Fig. 1, set 3). As soon as nitrogen supply is reduced (sets 1 and 2), the growth of the appearing leaves of the middle and upper storey is found to be inhibited, and they prove to be shorter than the leaves of such plants as received a liberal nitrogen diet (set 3). Starving plants show a rapid growth of the newly appearing leaves, as soon as they are given additional fertilizers, owing to which their young leaves become longer (set 4).

The ability of leaves rapidly to respond to a reduction or increase in the nitrogen diet received by the plant may be turned to advantage when eva-

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luating the requirement of wheat plants in additional nitrogen fertilizers. Thus, in our experiment with sand cultures, the plants of the first set were found to have their third leaves shorter than the second ones. This made us assume that an urgent need of nitrogen had appeared towards the beginning of growth of the third leaf, which coincided in time with the stretching of the point of growth. The same reasoning made us conclude that in the plants of the second experimental set nitrogen deficiency and expedience of additional nitrogen rations had coincided with the formation of the 4th leaf, or with the start of tillering.
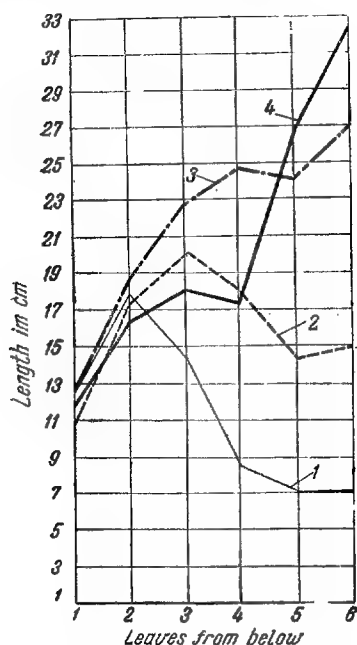

Fig. 1. Influence of nitrogen diet upon ultimate length of leaves in summer-wheat (sand cultures): *1*, Helriegel solution, minus N; *2*, same, plus 1/20 of normal dose of N; *3*, with a full normal dose of N; *4*, same, with 1/20 of normal dose when planted, 19/20 of the normal dose of N at tillering.
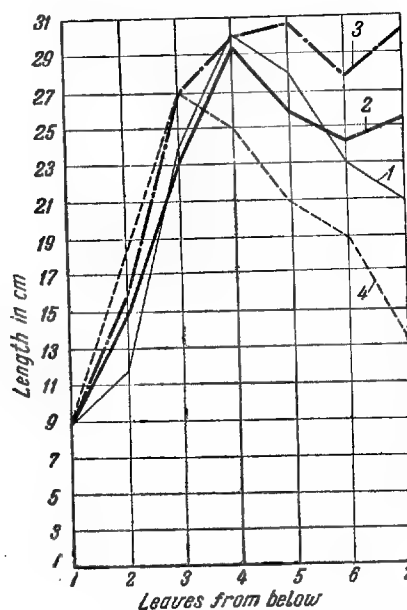
Fig. 2. Influence of nitrogen diet upon ultimate length of leaves in summer-wheat (soil-sand cultures). With a 90% moisture content: *1*, without fertilizers; *2*, K, P; *3*, K, P, N. With a 40% moisture content; *4*, K, P, N.

The correctness of our determination of starvation period in these plants was corroborated by the subsequent course of development and, in particular, by the development in them of ear rudiments. In the plants of the first set, in which nitrogen deficiency could be determined as early as at the beginning of the formation of the point of growth, when the formation of spikelets is at its initial stage, adult plants developed ears with a strongly reduced number of spikelets in them (5—7 spikelets per ear). The plants of the second set, in which the requirement for additional nitrogen diet became manifest later, developed correspondingly a much larger number of spikelets per ear than the plants described above (11—12). Lastly, in the plants of the third set, which were well-supplied with nitrogen, the number of spikelets per ear was greater still (14—16). Similar conclusions may be arrived at from the data of experiments carried out both with soil-sand cultures and in the field. In the former case (Fig. 2), owing to the content of a certain amount of nitrogen in the soil, the starvation of the plants which received no additional nitrogen diet (sets 1 and 2) and the necessity to supply them with

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such did not become evident until later on. Both proved to coincide with the start of growth of the leaves of the fifth storey.

Our investigation showed that, by measuring at regular time intervals the leaves that appear on plants which are known to have been well-supplied with nitrogen,and by comparing the values obtained with the results of leaf measurements in such plants as may be suspected of being deficient in nitrogen, it is possible to determine the beginning of nitrogen starvation, and, accordingly, to fix the dates of additional nitrogen rations to be administered. This can well be accomplished in spite of the delusive influence of other foreign agencies upon the growth of the leaves. However, if any of the growth factors is found to be badly deficient, as, for example, if the plant suffers a sharp water deficiency, then the diagnosis of nitrogen supply based upon the length of leaves becomes a poorly reliable one. The latter fact may well be illustrated by the data of measurements of leaf length in the plants of the fourth set, in which tillering plants were transferred to 40 per cent humidity (Fig. 2). The leaves of those plants showed a reduction in length, beginning with the fourth leaf, irrespective of the high nitrogen dose supplied. In this case the deleterious effect of water deficiency went to overlap the beneficial effect of liberal nitrogen supply.

Work on finding out adequate methods for diagnosticating at the laboratory nitrogen nutrition of wheat according to the external characters of the plant, is in progress.

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EXPERIMENTAL MORPHOLOGY

TRANSPLANTATION OF COMPONENTS OF OVOCYTE NUCLEI IN FERTILIZED NEWT EGGS

By G. V. LOPASHOV

Two questions appear to be of primary significance in the investigation of the problem as to the function of nucleus and cytoplasm in the development of multicellular organisms: 1) whether the determination of the cell types is bound with the differences in the functions of their nuclei acquired in the course of their development ([6],[8]), and 2) whether the formation of cytoplasmic zones differing in their developmental potencies ([7],[9],[10]) is bound with the distribution and action of the substances of nuclear sap issued by the giant nuclei of the ovocytes during the maturation of the latter ([1],[2],[5]).

The solution of a new problem—the problem of the mutual relations between the component parts of the cell—demands a new adequate method. Such a method seems to be, in distinction of the commonly used in studies on the physiology of development method of transplantation of groups of cells, that of intracellular transplantation, which has so far been applied by Hämmerling alone in his experiments on the unicellular alga *Acetabularia* ([8]). I was engaged in the working out of such a method for the purpose of studying the questions mentioned above in their application to newt eggs in 1940, 1941 and 1945. The present report is devoted to experiments of 1945, in which I attempted to investigate the action of the giant nuclei of the largest ovocytes and of their individual components upon the development of the fertilized eggs of *Triturus taeniatus* at the stage of 1—2 blastomeres by means of transplantation of the former into different zones of the latter.

The experiments were carried on under sterile conditions. Denudated eggs were operated upon and were kept on a bedding of a 2 per cent agar. The greater part of transplantations were achieved with the aid of thin glass pipettes, the pressure within which was raised by means of a specially designed screw syringe. To check the influence of the pricking itself on the development of the egg, about 400 experiments consisting only in pricking the eggs with glass needles in various directions were made (in 1940). The operations were made in Holtfreter's solution without $CaCl_2$ and $NaHCO_3$, saturated with cocaine (0.35 g NaCl + 0.005 g KCl + 0.028 g cocaine—100 cm³ of distilled water). In this medium the cytoplasm and the nuclear sap do not coagulate outside their membranes during a time necessary for the operation. After closure of the wound the eggs were transferred into Holtfreter's solution.

In the first group of experiments I introduced through a tear made in the surface coat entire nuclei of ovocytes into the thick of the cytoplasm of the eggs (by means of needles). The nuclei were soon expulsed. If I tore the nuclear membrane inside of the eggs, the nuclear sap remained to a certain extent within. It proved impossible to determine precisely the fate of the sap by such experiments.

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In the second group of experiments I used pipettes. To determine the fate of the sap I placed before the operation the isolated nuclei of the ovocytes for half an hour into a solution of vital stain (36 tests). The nuclei swelled, their diameter increasing about 1/3 of their original length, and stained slightly. We succeeded in developing a technique for sucking out of the nucleus only its sap without driving in any of the surrounding water. A small aperture was made in the surface coat of the egg by pricking it with a steel needle, the end of the pipette was inserted into this aperture and the sap was injected. If the opening closed at once, the sap invariably remained inside of the egg. The simultaneous introduction of the nuclear membranes produced no essential differences in the results; they are always very soon expulsed from the eggs.

The third group of experiments consisted in the injection of unstained sap sucked out of the nuclei immediately after their isolation from the ovo-cytes. With some experience it is possible to detect in which cases the sap exudes from the eggs in these experiments as well. I have distributed the operations into three classes: successful (+), dubious (?) and unsuccessful (—), according to whether the total sap or part of it remained within the egg and whether any saline had been introduced into the egg with it. Besides these experiments with sap injection, I put up 40 experiments with the introduction of cytoplasm or of saline.

The sap was introduced into the animal, marginal or vegetative zone of the egg (Table 1). In the latter two it was necessary to push the pipette somewhat deeper into the egg, and the injected sap is not visible.

Table 1

Degree of success	Transplantation of whole nuclei torn inside of egg	Injection of sap with pipette into different zones				Total pipette injections
		Animal zone	Marginal zone	Vegetative zone	Zone unknown	
+ . . .	—	20	5	6	5	36
? . . .	27	14	4	3	17	38
— . . .	5	4	5	3	14	26
Total	32	38	14	12	36	100

In the animal zone, in which the surface coat is very dense, it is possible by holding the pipette aslant to introduce sap exactly under the coat, where it is clearly visible as a vacuole. These vacuoles never break out to the surface of the egg. During several minutes they enlarge as if by jerks and finally merge with the cytoplasm. If stained sap is injected, the area of the egg's coat just above the vacuole becomes coloured. If the sap is injected into deeper parts of the egg, then the surface coat of the animal zone is stained. No coagulation phenomena which are characteristic to the action of water on the cytoplasm, occur upon the introduction of nuclear sap. They are not observed even after the introduction of stained sap containing a considerable amount of water. It may be assumed that when it enters into the composition of the nuclear sap the water is bound with the latter so intimately that it cannot do any direct harm to the cytoplasm. Injection of saline into eggs results in their disintegration: their surface coat is disrupted and the eggs flatten ou and are turned into small pools of cytoplasm.

The effect of pricking on the egg is in a noticeable number of cases its death. In my experiments of 1945, 28 eggs disintegrated before cleavage, 36 showed atypical cleavage and disintegrated gradually at the earlier stages, ending with that of the gastrula. A certain percentage of the embryos produced exogastrulae. In many cases various outgrowths were formed, however, a part of these must be put to the account of fold formation in the ectoderm where it was in contact with the agar surface. All these accessory phenomena were observed also in the pricking tests of 1940 and in the control tests with the transplantation of cytoplasm, while formation of folds and exogastru-

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lation were observed even in unoperated embryos developing without external membranes on agar since the beginning of cleavage. Because of insufficient sterilization of the agar 55 embroys were infected and developed only to the stage of the neurula. Consequently, only 27 of our test embryos can be taken into account when evaluating the effect produced by nuclear sap, and only 12 of them survived long enough to reach the later stages. Table 2 shows their distribution by types and by the results of the experiments (in brackets, folds from lying).

The ectodermal outgrowths were in most cases of an oblong shape and were situated in the dorsal part of the embryos. In several cases (not included in Table 2) they separated from the embryo as isolated ectodermal vesicles. These outgrowths contained no nervous or mesodermal cells. The cases denoted as enhanced gastrulation were distinguished by that the sinking of the entoderm inside proceeded at such an intense rate that it went on without the formation of the yolk-plug: the entodermal field sank into the embryo as a whole. In one embryo (N-182) a small «blastopore» appeared at the level of the middle of its body; later on it was displaced backwards, forming a secondary anal opening (sec. an.), near which arose a secondary «tail». Fig. 1 represents this embryo as a whole, Fig. 2, cross-sections through the area of the tails at the level of the secondary anal opening (A) and further back (B).

Table 2

Zone of injection	Results					
	Ectodermal outgrowths	Additional blastopores	Normal		Enhanced gastrulation	Exogastrulation
			straight	curved		
Undefined .	—	—	1	—	1	—
Animal . .	4 (1)	—	3	2	1	2
Marginal .	2	—	3	1	—	1
Vegetative .	1 (1)	1	1	—	—	1
Total	7 (2)	1	8	3	2	4

The secondary tail consists only of epithelium, mesenchyme and pigment cells, it contains no cells of the neural tube and axial mesoderm.

The number of experiments still remains quite insufficient. It may be assumed that all the deviations obtained were due to mechanical lesions provoked by the introduction of the sap. However, this must be considered less probable than the direct action of the sap, since: 1) the pricking of eggs and the injection of cytoplasm do not produce upon their development any effect leading to such results; 2) the sap after its introduction merges with the cytoplasm without creating any coagulated zones in it and in many cases does not alter the course of the egg's development.

From the above data a conclusion may be made that the changes of the local formative processes involve no changes in the histological destiny of the cells affected by them. This conclusion conforms with the fate of nuclear sap in normal development. At the maturation of the ovocyte the substance of the nuclear sap rushes towards the surface of the animal part of the egg where it enters into the composition of the surface coat which is forming at this time ([1],[2]). The differences in the density of this coat at the different surfaces of the embryo are one of the essential causes of the displacements, invaginations of its individual parts ([4]). Upon the introduction of additional sap local changes of the relative density of the surface coat may arise; it is, however, impossible to await regular and precisely localized changes in experiments like those reported above. Accordingly, the action of the sap might either provoke the phenomena described, or leave the course of the development unaffected.

The fact that the injected sap did not affect the histological destiny of the cells in the obtained outgrowths may be due to either of the two following circumstances. The histological destiny of the cells may be unconnected directly with the action of nuclear sap which is utilized for the formation of the latter; it may be the result of later interactions of the egg's parts.

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The fact that at the stage of 8 blastomeres the prospective mesoderm when isolated from the entoderm does not develop in its normal way ([10]) confirms this view. The other possibility is that in the course of maturation the sap diffusing within the egg participates in the determination of the local peculiar characteristics of the cytoplasm. Later on, however, it is already unable to influence its properties and affects only the density of the surface coat; these density changes can probably take place at the cost of nuclear substances supplied even at later stages of development ([2]).

Fig. 1.   Fig. 2A.   Fig. 2B.
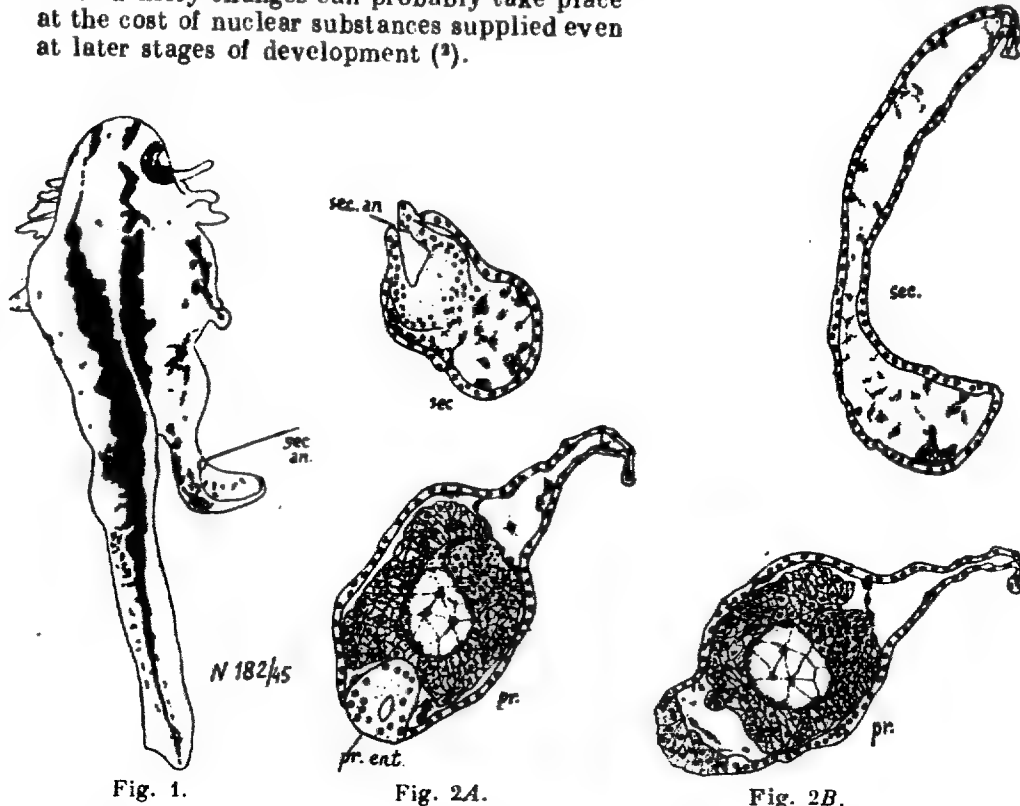
Main Conclusions. 1. Nuclear sap introduced into eggs of amphibians at the stage of 1—2 blastomeres does not cause death of the eggs or local coagulation of its endoplasm.

2. The introduction of nuclear sap into fertilized eggs either gives no effect on their development or manifests itself chiefly in the formation of ectodermal outgrowths.

3. No changes in the histological destiny of the cells of the outgrowths occur in this case.

4. The most probable explanation of these phenomena is that the injected sap in some cases changes the density of certain areas of the surface coat of the eggs, and this leads to the appearance of local formative motions.

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ZOOLOGY

BAIKALIAN MOLLUSCS FROM LAKE KOSSOGOL (MONGOLIA)

By M. KOZHOV

(Communicated by L. A. Orbeli, Member of the Academy, 9. III. 1946)

W. A. Lindholm described from lake Kossogol in 1929 (^{s}) a new molluscan species which he referred to the Baikalian subfamily *Benedictiinae—Kobeltocochlea michnoi*. The unique specimen of this species which had served as material for the description was afterwards presented to the Zoological Museum of the Academy of Sciences of the USSR, being lebelled «littoral fauna of lake Kossogol». This was the first finding of a representative of the native Baikalian fauna in lake Kossogol.

While collecting in 1944—1945 bottom fauna in the southern and northern parts of this lake, Anudarin discovered at a depth of 30—40 m one of the molluscan species from the Baikalian endemic genus *Choanomphalus*
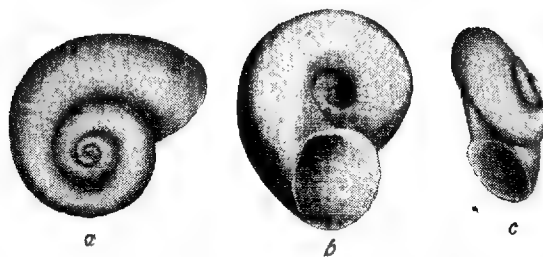

Fig. 1, *Choanomphalus mongolicus* n. sp. from lake Kossogol (Mongolia): *a*, viewed from above; *b*, from below; *c*, side-view. Magnification, 6 : 1.

(*Planorbidae*), which proved closely related to the Baikalian species *Choanomphalus amauronius* Bgt. Certain features which distinguish the Kossogol variety from the Baikalian species entitle us to raise the Kossogolian form to the rank of an independent species, for which I suggest the name of *Ch. mongolicus* n. sp.

D i a g n o s i s of *Choanomphalus mongolicus* n. sp. (Fig. 1). Shell small, top-like, compressed, colour horny-brown, surface finely striated, with deep rounded umbilicus, slightly upturned whorl and protuberant apex; forms 3—3.5 volutions. At the periphery the whorls are rounded, the last volution 2.5—3 times as wide as the last but one, near the aperture somewhat shifted downwards. Aperture round, with a slightly expressed anterior angle. Columellar edge, slightly turned outwards. Dimensions of the shell in mm: large diameter, 2.75—4.5; small diameter, 2.1—3.5; height, 1.4—1.75; number of volutions, 3—3.5. Locality: lake Kossogol (Mongolia), at a depth of 10—30 m.

The difference between *Choanomphalus mongolicus* and the Baikalian *Ch. amauronius* consists in that the former has a smaller and more thin-wal-

led shell (in *Ch. amauronius* the large diameter of the shell attains 6—8 mm) a less angular aperture and whorls rounder and increasing less abruptly.

In order to determine more exactly the relationship between the Kossogolian form and the Baikalian species of *Choanomphalus* and also with the closer forms of the genus *Planorbis*, the author investigated the radula and
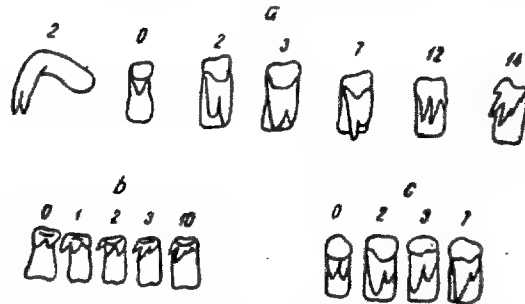

Fig. 2. Odontophores of radula: *a, Choanomphalus mongolicus* n. sp., magn. 840 : 1; *b, Planorbis gredleri* from environs of town Irkutsk, magn. 750 : 1; *c, Choanomphalus amauronius* Btg. from lake Baikal, magn. 750 : 1.

the sex apparatus of *Ch. mongolicus* from lake Kossogol, of *Ch. amauronius* from lake Baikal, and of a number of forms of *Planorbis (Gyraulus) gredleri* Gl. from the environs of town Irkutsk, from lake Baikal and lake Kossogol.
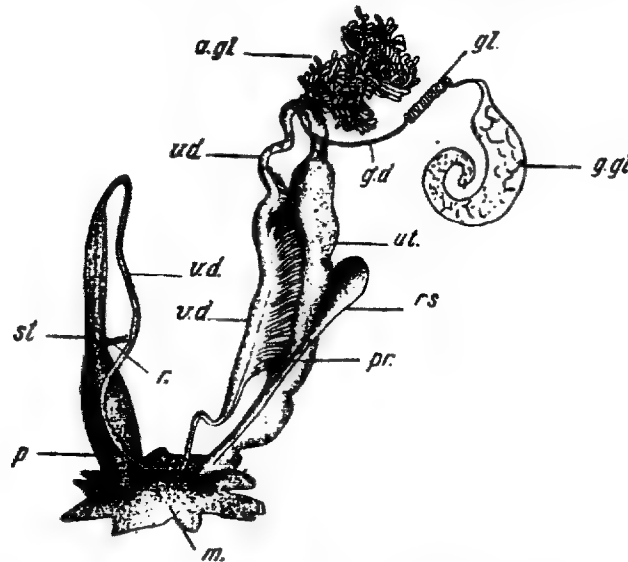
Fig. 3. Sex apparatus of *Choanomphalus mongolicus* from lake Kossogol: *g. gl.*, hermaphrodite gland; *gl.*, ball of hermaphrodite duct; *g. d.*, hermaphrodite duct; *a. gl.*, albumen gland; *r. d.*, seminal duct; *pr*, prostate gland; *p*, penis; *st.*, penis stylet; *r.*, penis muscle-retractor; *m.*, muscles around sex openings; *ut.*, uterus; *rs.*, receptaculum semini.

A comparative study went to demonstrate that the Mongolian forms described above should be referred to the genus *Choanomphalus*. In *Ch. amauronius* from lake Baikal and in *Ch. mongolicus* the incisors of the odontophores of the radula, from the 1st to the 7th—8th counting from the middle, bear two tooth-like processes (Fig. 2), whereas in *Pl. gredleri* they are armed with three teeth each. The hermaphroditic gland in *Pl. gredleri* is somewhat larger

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and with larger lobules than in *Choanomphalus* (Fig. 3). The ball of the herma-phrodite duct in the former is considerably larger than in *Choanomphalus*; the prostata is also more vigorously developed. In *Pl. gredleri* the seminal duct in its extended distal part, before it enters the penis, shows a constriction, while no such constriction is present in *Choanomphalus*. In *P. gredleri* the loop of the seminal duct between the prostata and the penis is partly embedded in the muscles connecting the ends of the female and male sex ducts; in *Choanomphalus* it lies open to view. The length of the spermatheca of *Planorbis gredleri* is nearly a half that of *Choanomphalus*. No essential differences between the sex apparatus in *Ch. amauronius* from lake Baikal and *Ch. mongolicus* from lake Kossogol could be established.

Lake Kossogol is situated west of the southern end of lake Baikal, at a distance of 245 km from it, along the aerial line. Its absolute altitude is 1640 m, which means that lake Kossogol is situated at an altitude surpassing that of lake Baikal by about 1200 m. The area occupied by the lake is 2612 km^2; its maximum depth, about 250 m. From the west the lake is bordered by the mountain range Ulan-Taiga (absolute altitude above 2000 m); from the north, the mountain knot Munku-Sardyk with an absolute height of 3491 m. A small glacier is gliding down the summit of Munku-Sardyk. The southern part of the lake gives rise to the river Eghingol discharging into the Selenga River, which is a tributary of lake Baikal. The width of this river in its upper reaches is no more than 4 m, its depth attaining up to 1—1.5 m.

The following two alternatives may be suggested to account for the occurrence in lake Kossogol of representatives of a fauna endemic of lake Baikal.

1. In its past, lake Kossogol was directly connected with lake Baikal, forming a part of the latter, like the extinct ancient lake Tunkin whose ample depression is situated between lake Baikal and lake Kossogol. If this was not the case, the lakes might, like many other relic lakes, enter into the composition of the ancient Baikalian group of lakes where developed fauna of the Baikalian type ([2],[4]). During the Quaternary those lakes became separated from one another by the general uplift of the area. Thus, the Baikalian molluscs dwelling in lake Kossogol are relics of the fauna of the ancient Baikal, in the same way as *Polychaeta* is a relic of Baikalian fauna in the Zipo-zipikan lakes, situated in the basin of the middle course of the Vitim River ([3]).

2. In the past lake Kossogol formed one of the links of the channel through which the ancient (Tertiary) Central-Asiatic hydrofauna penetrated into lake Baikal. Some of the representatives of this fauna have survived in lake Kossogol; however, they came to thrive in lake Baikal only. At this juncture it seems natural to assume that lake Kossogol had been connected with lake Baikal by a river more abundant in water that the recent Eghingol is.

The latter alternative seemed to me to be more probable, which I declared in my paper on the molluscs of lake Baikal ([2]). This view was likewise advocated by Vereschaghin ([1]). However, the question cannot be settled in a definite manner until the hydrofauna of Mongolia and of the adjoining regions of South Siberia is investigated in more detail. In the various parts of Mongolia, especially in the north-west of that country, there occur numerous large and deep fresh-water and brackish lakes whose hydrofauna is utterly unexplored. It is not improbable that upon a thorough investigation of those lakes they will be found to harbour some remnants of the «Baikalian» fauna, in the same way as the sponge *Baicalospongia* had been found in lake Jeghetay-Kul in Urenkhay.

On account of the discovery in lake Kossogol of a representative of the Baikalian genus *Kobeltocochlea*, it was pointed out by Lindholm ([5]) as far back as in 1929 that the Baikalian subprovince of the Holarctic as defined for lake Baikal by L. S. Berg, should be made to comprise also lake Kossogol. Now that this lake has become the scene of new discoveries, the above opinion appears to be based on still more solid foundation. But aside from lake Kossogol, to be included into the Baikalian Holarctic subprovince deserve

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also lakes of Zipo-zipikan (the basin of the middle course of the Vitim River). These lakes are remnants of the ancient Baikal, and they do likewise contain certain elements of the Baikalian fauna, such as the *Polychaeta Manayunkia baicalensis* Nusb. ([2]). and the *Gobius* of the Baikalian genus *Limnocottus*, which were discovered there by an expedition sent out by this Institute in 1939—1941.

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ZOOLOGY

ON PURELY MATROCLINOUS PROGENY RESULTING FROM CROSSES BETWEEN CERTAIN TELEOSTEAN FISHES

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It has been established that in crosses between very remote forms, belonging to different classes or even types of invertebrates, the spermatozoon only activates the egg without taking part in its development. This phenomenon has been given the name of gynogenesis *. Purely matroclinous pseudohybrids obtained as a result of such crosses show usually a pathological character and perish during the period of embryonic development.

It is much more difficult to decide whether we deal with true or false hybridization in the case of crosses between less remote forms, in particular between species belonging to different orders, suborders or families of fishes. Loeb [2] and Newman [3] have arrived at conflicting conclusions in this question. Loeb considers the progeny of such «heterogeneous» crosses as a purely matroclinous one and, consequently, parthenogenetic; Newman, finding in them paternal characteristics, regards them as true hybrids. The paternal influence has been reported for inter-order crosses of teleostean fishes also by Kryzhanovsky [4], but, on the other hand, this author observed also parthenogenctic development in inter-familial crosses of ♂ *Schizothorax pseudaksaiensis* × ♀ *Diplophysa dorsalis* and in reciprocal inter-order crosses *Leuciscus schmidti* × *Diptychus dybowskii*.

In 1932 C. and L. Hubbs [5] published their report on gynogenesis established by them in natural conditions in one of the forms of the family *Poecilliidae*, the *Mollienisia formosa*. The latter was earlier described as an independent species. C. and L. Hubbs regard it to be a hybrid between *Mollienisia sphenops* and *M. latipinna*. Considering that this hybrid produced a purely maternal progeny composed only of females, these authors conjecture that in this case the activation of its eggs was achieved by spermatozoa of one of the parent species. Howell [6] and Meyer [7] find the arguments brought in support of this hypothesis inadequate, and the existence of gynogenesis in vertebrates under natural conditions, that would yield a progeny able to attain maturity, highly improbable.

We arranged 21 crosses between species belonging to different orders (*Cypriniformes, Perciformes* and *Esociformes*) of teleostean fishes with the following results. The embryonic development was, as a rule, limited to the very earliest stages (cleavage, gastrulation, beginning of differentiation of the body of the embryo); on most rare occasions a few unviable larvae were hatched. In none of the experiments could we detect in the progeny any paternal characters, as we have already reported in an earlier paper [8].

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* Most authors regard gynogenesis as a particular case of parthenogenesis. Wilson [1], however, considers that gynogenesis does not fit into the conception of parthenogenesis, since in the latter the activation of the egg is due to agencies other than the spermatozoon.

In 1938 we studied in detail the progeny of the inter-order cross ♀ *Scardinius erythrophthalmus* × ♂ *Esox lucius*. The greater part of the spawn developed abnormally and perished in the course of the first 24 hours upon fertilization. A considerable part of the spawn perished also during the period of gastrulation and of the initiation of the embryonic body. Only a few embryos attained the stage of hatching. The larvae which developed a few days after they were hatched showed, as a rule, a more or less pathological aspect. Only two of them seemed normal and swam with ease. In these two one could observe not only heart pulsation but also the circulation of blood in the vessels, which could not be seen in the rest of the larvae. In the distribution of their pigment cells (in 3 rows), in the number of their myotoms (37—38) and in the general aspect these two larvae differed in nothing from the control larvae of *Scardinius erythrophthalmus*. The other larvae also showed a general similarity to the control ones, differing from them only by growth anomalies. In none of them could we detect paternal characters (*i. e.* those of *Esox lucius*). Therefore one is naturally induced to think of their gynogenetic origin. Of course, this question cannot be definitely solved unless on the ground of cytologic investigations.

The possibility is not to be excluded that the progeny obtained from the same cross may include both true and false hybrids. At least in some crosses a heterogeneity of the progeny, presenting certain interest in this respect, was observed: in the more successfully developing individuals the paternal influence was found to be suppressed, while the maternal one was predominant, or *vice versa*. Newman (*), it is true, considers even pure matroclinous larvae to be true hybrids and believes that the paternal hereditary factors are possessed by them in a latent state, but he does not provide any proofs to support this thesis. We admit the possibility of gynogenesis in teleostean fishes also upon the ground that the eggs of some of these (*Acerina cernua*, *Rutilus rutilus*) can develop even without any fertilization and produce parthenogenetic embryos, or even larvae (*).

Considerable interest belongs in this respect to certain inter-order crosses between female *Carassius carassius* and particularly, the cross ♀ *Carassius carassius* × ♂ *Leuciscus cephalus*. This cross was achieved by us three times, the results being roughly similar in every case. A considerable part of the spawn perished at the very beginning of development. Most developing embryos and larvae were pathological to some extent. In some of them paternal characters could be detected, distinguishing them from the respective control embryos and larvae of *Carassius carassius* (an increased—up to 15—number of caudal segments, *etc.*). The few normally developed larvae were absolutely similar to the controls: they showed a similar short caudal section of the body, the same number of segments (20—22 thoracal and 10—12 caudal), a similar pigmentation, shape of the yolk, and a powerfully developed *vena subintestinalis* on the surface of the yolk, so typical of *Carassius carassius* larvae.

In two of the experiments we were able to obtain but a few young of a pure *Carassius carassius* type. In 1941 this cross permitted to breed 49 fully viable young which grew up to the size of 4—5 cm. All these showed also a pure matroclinous type. In every single young fish we determined the number of rays in their dorsal and anal fins, and in 10 of them a number of other quantitative indices. In the table we compare the indices of the investigated characters of hybrid young obtained from the cross ♀ *Carassius carassius* × ♂ *Leuciscus cephalus* with the average figures for 25 individuals of *Carassius carassius* from the Medovo Lake in the flood-plain of the upper stream of River Don (the female *Carassius carassius* used for the cross with *Leuciscus cephalus* had been caught also in this lake) and with the figures for *Leuciscus cephalus*, according to Berg ([10]).

The indices found for the cross ♀ *Carassius carassius* × ♂ *Leuciscus cephalus* concord well with Berg's formula for *Carassius carassius*: DIII—IV

14—21, AII—III 6—8, 1.1. $28\frac{6-8}{(5)\ 6-7}$ 37, gill-rakers 23—33. The number
of vertebrae which is not indicated by Berg for *Carassius carassius* or *Leuciscus cephalus* is in the young of their cross within the range of variation observed in the *Carassius carassius* of Medovo Lake. In *Leuciscus cephalus* the number of vertebrae is considerably higher (no less than 42, according to our findings

Characters	♀ *Carassius carassius* × ♂ *Leuciscus cephalus*		*Carassius carassius* (from Medovo Lake)		*Leuciscus cephalus* (according to Berg)
	range of variation	M	range of variation	M	
Number of rays in dorsal fin	III—IV 14—19	16.7	III—IV 16—19	17.2	III 8
Number of rays in anal fin	III 6 (7)	6.0	III (5) 6	6	III (8) 9 (10)
Number of vertebrae	32—34	32.4	31—34	32.5	—
Number of gill-rakers	25—29	28.1	28—33	30.4	8—10
Number of transversal rows of scales	32—34	33.1	32—35	32.9	43—47
Number of scales above l. l.	7—8	7—8	7—8	7.0	7—75
Number of scales below l. l.	(6) 7	6.9	(6) 7	7.9	3
Formula of pharyngeal teeth	4—4	—	4—4	—	2.5—5.2

on a few individuals). As one gathers from the table, the young from our cross show the greatest difference from those of *Carassius carassius* of Medovo Lake in regard to the number of rays in the dorsal fin and to that of gill-rakers, but any essential significance should hardly be attributed to these differences (it is possible that the formation of gill-rakers in the investigated young was not yet complete).

Thus, in all the three crosses only an insignificant part of the progeny proved viable, this part was of a purely matroclinous type; the rest developed abnormally and perished before hatching or shortly after it. The question naturally arose, might it not be that this progeny was composed of two genetically different groups of individuals: one, consisting of pathological forms, being composed by true hybrids; another, consisting of normal forms,—by false (gynogenetic) hybrids?

Results similar to those described above were obtained by crossing ♀ *Carassius carassius* × ♂ *Scardinius erythrophthalmus*. The few larvae developed were all of a purely matroclinous type. Only 4 individuals reached the stage of the young fish. The following formulae for dorsal and anal fin rays were obtained for them: 1) DIII 14, AIII 7; 2) DIV 16, AIII 7; 3) DIII 17, AIII 6; 4) DIV 17, AIII 6. These formulae do not pass the limits of the range of variation observed in regard to these characters in pure *Carassius carassius*.

The crosses ♀ *Carassius carassius* × ♂ *Gobio gobio* or *Wimba wimba* gave similar results: in both cases the progeny obtained showed, at least at

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the larval state, pronounced—if not absolute—matrocliny. We did not succeed in bringing them up to the young fish stage.

From the point of view of the questions raised in the present report it should be of interest to subject to a special investigation also *Carassius auratus gibelio*, in particular, in connexion with the unanimous statements of the fish-breeders of some parts of the Ukraine that in certain fish ponds of these parts, *e. g.* in the fish-breeding farm «Rotok» (Kiev-province), the males of this *Carassius* species are, as a rule, absent. This circumstance has been corroborated by a limited number of our personal observations.

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